PROCEEDINGS

OF THE

AMERICAN PHILOSOPHICAL SOCIETY

HELD AT PHILADELPHIA

FOR PROMOTING USEFUL KNOWLEDGE

VOL. XLVIII APRIL–AUGUST, 1909 No. 192.

THE EVOLUTION OF THE CITY OF ROME FROM ITS ORIGIN TO THE GALLIC CATASTROPHE.

BY JESSE BENEDICT CARTER.

(*Read April 22, 1909.*)

In a normally constituted man time and space are in permanent coördination. In the world of historical science such a permanent coördination is sought after, but not yet everywhere obtained. The student of history and the student of topography are too apt to work in ignorance of each other. The history of Rome has usually been written with small regard for that material and physical thing, the city of Rome; while the writer on topography is far too apt to see the buildings and the piazzas of ancient Rome as an empty stage, a place for action, but for an action in which he is not professionally interested.

Yet the transition through which so many of the natural sciences have recently gone, the change from being merely descriptive to being biogenetic, ought to serve as a lesson to the topographer. It is not possible to study even the site of ancient Rome without taking into account the vicissitudes of history in which this site has been involved.

I would accordingly ask your attention today to an attempt to sketch in its outlines the development of the city of Rome from its earliest beginnings through the Gallic catastrophe. Such a bio-

PROC. AMER. PHIL. SOC., XLVIII. 192 J, PRINTED SEPTEMBER 2, 1909.

graphical sketch (for under this treatment the city itself becomes endowed with life and the product is veritably a biography) covers a distinct field in that long series of periods which follow one another in the story of the Eternal City.

Yet this period of the origins has been strangely neglected by modern scholars, at least in so far as attempts at the coördination of material are concerned. The student of ethnography has formed his own opinions regarding the early settlement of this part of Italy, the student of language has drawn his own deductions; the student of religion has discovered certain perfectly definite things regarding the civilization of these primitive peoples; and the student of topography has made his own discoveries, but has also held his own counsel. Yet the language of communication between these special students has been in the main the old traditional one of Rome's founding.

The greatest difficulty which confronts the student of the origins of Rome is not the absence of statements regarding it, but rather the superabundant presence of such statements. If what was afterwards the great city of Rome had been entirely unknown in its birth, we would have placed it in the category of many other famous individuals, and thought nothing of it. But the presence of such a plenitude of sources has at least two bad results; first it leads to endless and hopeless attempts to reconcile conflicting statements[1]; and second even after our reason has convinced us that these statements are without authority and represent merely the late products of artificial legend making, we have great difficulty in casting them to one side, and we unconsciously and instinctively recur to them, so much are they a portion of our intellectual heritage. We may prove that Romulus was not known in Rome until after the Gallic catastrophe,[2] and that we have no reason to suppose the Palatine settlement to be any older that the Capitoline or the

[1] Compare the attempts periodically made to reconstruct the early history of Rome on the basis of the legendary accounts.

[2] See Carter: "The Death of Romulus," *American Journal of Archæology*, 1909, pp. 19–29; and (more fully) my forthcoming article, s. v. "Romulus," in Roscher's "Lexikon der griechischen und romischen Mythologie."

Quirinal,[3] but out of the ruins of our tradition Romulus, Remus and the wolf arise. Thus it is that we are still presenting the subject according to the scheme and phraseology of Varro, though there is scarcely any other part of Varro's learning which we accept unhesitatingly.

In the first place our study of Roman religion and its coördination with the study of the primitive religions of today have shown us that, down to the dawn of history, the inhabitants of the region of Rome were a semi-barbarous people. Their religion was still involved in animism. They felt themselves surrounded by a countless host of potentialities, whose names they knew, but of whose nature they were otherwise ignorant, except in so far as that nature externalized itself in definite acts.[4] Their religious organization shows that this primitive people was divided, as its most original division, into curiæ or brotherhoods, and that every member of the community must of necessity belong to one of these curiæ.[5] Their religion shows us further that their interests were agricultural.[6]

Further we know that they lived in little communities on the hilltops surrounded by a circular wall or stockade. Such a primitive settlement was certainly not a city—an urbs. At best it might be dignified by being called a town, an oppidum.[7]

The geological character of the campagna, the presence of vast

[3] See below, and also "Roma Quadrata and Septimontium," *American Journal of Archæology*, 1908, p. 181.

[4] See Wissowa: "Religion und Kultus der Roemer," p. 20, "Sämmtliche Gottheiten sind sozusagen rein praktisch gedacht als wirksam in all denjenigen Dingen, mit denen der Roemer im Gange des gewöhnlichen Lebens zu thun hat"; and Carter, "Religion of Numa," p. 5 ff.

[5] If we accept the theory that matriarchy existed in Rome before the institution of the patriarchal system, we are virtually driven to consider the Curiæ as preceding the family. For an excellent discussion of the Curiæ, cp. Eduard Meyer, "Geschichte des Altertums," Vol. II., p. 511 ff.

[6] Cp. the table of gods for this early period, as reconstructed by Mommsen, "Corpus Inscriptionum Latinarum," Vol. I., Part 1, ed 2, p. 288, or by Wissowa, "Religion und Kultus," p. 18 and cp. p. 20: "es spiegeln sich in ihr (der alten Götterordnung) die Interessen einer in Ackerbau und Viehzucht . . . lebenden Gemeinde."

[7] Cp. the investigations of E. Kornemann, "Polis und Urbs," in "Klio Beiträge zur alten Geschichte," 1905, p. 72 ff.

quantities of running water, and the consequent erosion, produced a large number of tongue-shaped or circular elevations, admirably suited to such settlements.[8] These clusters of houses surrounded by a ring-wall were merely habitations. The people tilled the fields in the valleys below. It is impossible for us to distinguish clearly between these hill-top towns in their early history. They were probably very similar in population and consequently in customs. Judging however by the presence or absence in historic times of old cult centers it would seem that there was no settlement upon the Aventine,[9] possibly because it was too close to the river. Nor does there seem to be any particular justification for supposing that the Palatine was in any sense the leader in this group of hill towns, by virtue either of its superior age or of its greater influence. The Palatine is singularly free from old cult associations.[10] Such associations as seem old are connected with the later legends, for example that of Romulus and Remus, which did not arise until the fourth century, and even in these cases the Capitoline offers a distinct rivalry to the Palatine.[11] It is easy to understand how at a later day the Palatine might have been elevated into this position of superiority.[12]

[8] Cp. the presentation of Richter: "Topographie der Stadt Rom," p. 25, 26.

[9] At least in later times it is known as pagus Aventinensis, CIL., XIV., 2105 (inscription from Lanuvium); and the fact that it was later opened to the plebeians for settlement would indicate the absence of any older settlement. The town of Aventum is an unfortunate suggestion of Jordan ("Topographie," I., 1, 182) and never had existence. Cp. Huelsen in Pauly-Wissowa's "Encyclopædie der classischen Altertumswissenschaft," s. v Aventinus, Sp. 2283, 23 ff.

[10] Cacus and the very doubtful Caca, in whom Wissowa ("R. und K.," p. 24, note 1) is inclined to see a pair of ancient gods, belong really on the Aventine rather than on the Palatine. Huelsen's statement (Jordan-Huelsen, I. 3, p. 45), "von den Kulten auf dem Palatin cheinen einige in sehr alte Zeit hinauf zu gehen, wie der der Febris, der Fortuna, der Dea Viriplaca, der Luna Noctiluca," must be taken merely relatively, as none of the deities mentioned (with the exception of the uncertain Dea Viriplaca) precede the later kingdom.

[11] Cp. the rival casa Romuli on the Capitoline; and the Salii Palatini versus the Salii Collini.

[12] Owing to its popularity as a residence during the closing years of the Republic, and the preference of Augustus and his successors.

This little group of towns is not as yet however the city of Rome: it is possible that in the course of time it might have become the city of Rome, either by the superior power of one oppidum which would shortly have added the others to its territory, in somewhat the way in which the traditional account considers that Rome was actually founded,—the Varronian scheme, which proceeds from the presupposition of the primacy of the Palatine,—or by some sort of reciprocity, resulting in union, of which we see the first traces in the annual joint sacrifices of the Septimontium.[13] But either one of these ways would have required a very long period of time, and in either case the intellectual development of the people would have been continuous so that the traces of barbarism even in the conservative field of religion would have been much fewer in number. Every indication points to a rapid change and one which affected the towns equally. Such a change could come only from outside, and from a people superior to Rome in culture. When we ask what this people was, the answer comes more clearly every year,—the Etruscans.

It seems fairly certain that the Etruscans as we know them in the history of Italy were a composite people made up of a native Italic stock combined with an invading stock, whose original home was in Asia Minor.[14] Further it seems probable that the invading stock came by sea across the Mediterranean and landed on the western coast of Italy, and that their advent did not precede the beginning of the eighth century.[15] Allowing them about two centuries

[13] On the Septimontium, compare Varro, L. L. 6, 24: dies Septimontium nominatus ab his septem montibus, in quis sita urbs est, feriæ non populi sed montanorum modo, ut paganalia qui sunt alicuius pagi; and the interesting treatment by Wissowa in the Satura Viadrina-Gesammelte Abhandlungen, p. 230 ff. Cp. also Platner: "Classical Philology," I., 1906, p. 69.

[14] The hypothesis of the East, more especially of Asia Minor, as the original home of the Etruscans is at present pretty generally adopted. Their acquaintance with the Babylonian haruspicina and with Greek mythology, the general plan of their houses and the shape of their helmets all indicate an eastern origin. For details see the admirable résumé of the present condition of the Etruscan problem by Körte in Pauly-Wissowa s. v. Etrusker.

[15] Whether the Etruscans came by land or by sea is still a subject of discussion, though the hypothesis of the sea route seems to be gaining strength at the expense of the other. There seem to be traces of their movement on

to accomplish their amalgamation and conquer the region afterwards known as Etruria, they would come into contact with the Roman stock in the plain of Latium about the beginning of the sixth century.[16]

The Etruscans, therefore, a sea-faring and so a city-loving folk, conquered these hill towns and enclosed them all together with the intervening valleys with one wall. But before building this wall, they drew the plough about the space to be enclosed and thus created the pomerium *ritu Etrusco*.[17] We do not know very much about their wall but we do know about the pomerium, and as the wall was surely inside of it,[18] we have a general idea of its position.

the islands of the eastern Mediterranean, especially on Lemnos, where an inscription practically Etruscan in character has been found. It is uncertain exactly what we are to call these people before the "Etruscan" people were brought into being by the amalgamation of this immigrant stock with the Italic stock. It has been suggested with a reasonable degree of probability that they were the Pelasgians. The date at which they entered Italy is a matter of some considerable uncertainty. The date as given above (circa 800) depends upon the validity of the supposition that in the long series of tombs which the cemeteries (especially near Bologna) show, the earlier tombs are not of the Etruscans but only the later ones, the tombe-a-corridoio, and the tombe-a-camera. However several scholars, who are in hearty accord with the eastern origin, and the journey by sea, are not content with so late a date as the eighth century, on the ground that it does not allow sufficient time for the development of the Etruscans in the peninsula of Italy. According to them the coming of the Etruscans should be placed two or three centuries earlier.

[16] This date corresponds with the tradition of the later kingdom. Tarquinius Priscus reigned thirty-eight years, Servius Tullius forty-four years, Tarquinius Superbus twenty-five years, a total of one hundred and seven years, which added to B. C. 509, the supposed year of the founding of the Republic, gives B. C. 616, as the beginning of the so-called Later Kingdom. Such an agreement may be of absolutely no value, on the other hand it may have a certain significance if the tradition represents the faint reflection of the period of time when the new influence came.

[17] Not only the Pomerium, but the whole idea of delimitation seems to have come to Rome from Etruria. Much of the terminology of Roman surveying bears the imprint of Etruria. Roman tradition recognized the Etruscan origin of the Pomerium: cp. Varro, L. L. V., 143: oppida condebant in Latio Etrusco ritu multi, id est iunctis bobus, tauro et vacca, interiore aratro circumagebant sulcum.

[18] On the whole question of the pomerium and its relation to the city wall, compare *American Journal of Archæology*, 1908, p. 177.

Thus was created what the topographers call "the city of the four regions."[19] It would be preferable to use the old Roman term *urbs et capitolium,* for this city, the *urbs* did indeed contain four regions, but apart from the city though inclosed in the same wall was the citadel, the *capitolium.*[20] Such an arrangement is in itself an added proof that the Palatine was not the ruling spirit. The Etruscans coming from without were free from prejudice and chose the Capitoline as their citadel simply because it offered superior advantages from the fortificatory standpoint.

On the Capitoline arose the Etruscan temple of Jupiter, Juno and Minerva. It is strange that the Etruscan character of this cult has not been more readily recognized. Minerva herself is more than half an Etruscan deity, hitherto unknown to Rome,[21] and the triad, Jupiter-Juno-Minerva, is a favorite among the Etruscans. The temple was built in the Etruscan style by Etruscan workmen and the ornamentation and the very images of the gods came from Etruria.[22]

With the coming of the Etruscans begins a tradition which has in part an historical value. This tradition presents us with the figure of Servius Tullius, unquestionably a real person, probably the

[19] "Die Vierregionenstadt" of the Germans. I do not know of any instances of the term in antiquity. The ancient term seems to have been urbs et capitolium.

[20] The capitolium had of course a protecting wall of its own. This is clear from the fact that it was capable of being held against the Gauls, even after the Gauls had captured the city proper. The other hill-top oppida which were included in the urbs certainly had walls of their own, but these walls probably ceased to be kept up after the large surrounding wall was built. In the case of the Capitolium however the original wall was preserved and probably strengthened.

[21] Minerva has no festival in the old calendar, the so-called calendar of Numa. The Quinquatrus which occurs in that calendar and which is ordinarily associated with Minerva had originally no connection with her, but belonged entirely to Mars. Minerva's cult seems to have originated at Falerii and to have spread from there into Etruria and also into Rome. On Minerva, cp. Wissowa in Roscher's Lexikon, s. v. Minerva, and "Religion und Kultus," p. 203; and Carter, "Religion of Numa," p. 44 ff.

[22] The image of Jupiter came from Etruria; compare Pliny (N. H., XXXV., 157) and Ovid (F., I., 201 ff.); also the quadriga on the roof (Pliny, *l. c.*). The workmen employed on the temple gave the name to the Vicus Tuscus, where they lived.

first historical character in the annals of Rome. But though the character of Servius is a real one, legend has added many of the "events" attributed to him. One of these events concerns our own theme—it is the building of the wall of Rome. The tourist knows this wall as the inner of the two walls, of which traces still remain in Rome, that wall of which there are remnants beside the railway station and on the Via Nazionale.[23] Up to the present the statement that Servius built a wall has been accepted as an historical fact, and though it was recognized that the so-called Servian wall as we know it dates from the end of the fourth century before Christ, scholars have almost always assumed that there was another wall on the same spot and that this previous wall dated from the Servian age.[24] But, as I hope to be able to show in a moment, this is an altogether gratuitous assumption, and serves simply to hinder the understanding of history. In the first place there is absolutely no proof that Servius Tullius built a wall, other than the name "Servian wall" which attaches to a structure obviously of the fourth century. The tradition would in any case be worthless, but we have not even a consistent tradition. A study of the growth of the city as attributed to the various kings brings no profit, but exhibits merely a mass of contradictions and inconsistencies.[25] So far as the name

[23] Sections of this wall are constantly being discovered. At the date of writing (April, 1909) a very fine piece has been unearthed near the Spithoever property.

[24] The only exception to this statement known to me is Eduard Meyer (Hermes, XXX., 1895, p. 13): "dass die Servianische Mauer nicht älter ist als das vierte Jahrhundert, ist seit O. Richter's Nachweis unumstösslich. Sie umschliesst die Grossstadt der Samniterkriege." That this statement has not been more appreciated is doubtless owing to the fact that it is capable of being understood to apply merely to the date of the actually existing Servian wall, leaving always the possibility that it implies another wall on the same site preceding the "Servian" wall.

[25] In Dionysius of Halicarnassus (4, 13) and in Strabo (p. 234M) Servius Tullius is aid to have added the Esquiline and the Viminal; but Livy (1, 44, cp. the author of de vir. ill. 7) says that he added the Quirinal and the Viminal and increased the Esquiline; whereas the Quirinal is elsewhere (Dionys. 2, 50, Strabo, p. 234M) supposed to have been included in the city of Romulus and Titus Tatius. On the other hand the so-called Servian wall included the Aventine, hence Servius is supposed to have added this hill to the city, whereas a very strong ancient tradition attributed the

itself is concerned, in the minds of the contemporaries and successors of Cato a wall at that time nearly two hundred years old would be easily associated with the kingdom and might readily be named after the most famous of the kings, Servius Tullius. There are in other words no traces of a real Servian wall either preserved in monumental form for the topographer or found in the historical records. The occasional references found in Livy to the gates of what we know as the "Servian Wall," in connection with events which happened at or before the Gallic catastrophe, are most rightly explained as anachronisms, and they offer no difficulty to one who is accustomed to the vagaries of the Roman historians.[26]

On the contrary, it is on the face of it extremely unlikely that an enlargement of the city limits would have been necessary so soon after the building of the large encircling wall which we attribute to the Etruscans. Yet, as a matter of fact, the so-called "Servian Wall" includes a much larger space than the wall of the "Four-Region City."[27] It includes on the northeast the high tableland where the Quirinal and the Viminal unite, but still more important it includes the Aventine. It is the inclusion of the Aventine which creates the chief difficulties in understanding the history of Rome until after the Gallic catastrophe. Let us try the experiment of considering the Aventine as a suburb and of reading our history under such a condition.[28] The city which the Etruscans founded and in which Servius Tullius lived, and according to our present assumption the only city of Rome until after the Gallic

addition to Ancus Martius (Cicero de rep. 2, 18; Dionys. Hal. 3, 43; Strabo, p. 234M; Liv. 1, 33; de vir. ill. 5). The difference of opinion regarding the Caelian is still more marked. On the whole question compare Jordan, "Topographie," II., p. 206, 207.

[26] *E. g.*, Livy (5, 41) speaks of the Gauls as entering by the Porta Collina, referring doubtless to the gate in the "Servian" wall, as it existed in his day.

[27] At this point the reader may be inclined to challenge these statements and to ask what we know of the course of the wall of the Four Region City. Of the wall itself we know nothing, but we do know that it lay inside the pomerium, and we know approximately the course of the pomerium, and to what extent it in its turn lay inside the Servian wall.

[28] It may require a certain amount of practice to conduct this experiment successfully, just as it takes practice to eliminate the arch of Severus in reconstructing the Forum of the Republic and early empire.

catastrophe, was that particular form of the city which the topographers call "the city of four regions" and which was more familiarly known in history as *urbs et capitolium*.

In the first place we note the permanency of the phrase *urbs et capitolium*[29] and we ask whether it is likely that the phrase would have obtained such immortality if the form of the city to which it was applicable had so soon given way to the other form, the so-called Servian city. The permanence of the name seems to argue for the long existence of that particular city from which the name was derived. In the second place the annals of religion offer us in this early period at least this knowledge, namely, the establishment of temples to various deities more or less strangers to Rome, in the region outside of the pomerium.[30] One of the most important of these deities was Diana. She came into the religious life of the state merely because of her connection with the Latin league, and her temple was not a temple of Rome alone but of the whole league.[31] This temple was situated on the Aventine,[32] and while of course it was outside the pomerium it has always been difficult to understand why Rome made bold to put a league temple inside her city wall, when all the expanse of the Campus Martius was at her disposal. But if as we are now supposing the Aventine also was a suburb, the difficulty disappears. Conversely when the temple of Apollo[33] was built, while it must of necessity have been outside the pomerium, it is difficult to see why it should have been placed in the exposed Campius Martius, when there was the possibility of placing it on the Aventine itself outside the pomerium but sup-

[29] Urbs et capitolium occurs; Cæsar de bell. civ. 1, 6, 7; Liv. 3, 18, 6; cp. Liv. 38, 51, 13; Flor. Epit. 2, 6, 45; Jord. Rom. 202.

[30] A useful list of these temples and their dates is given in Wissowa's "Religion und Kultus," p. 516 ff. It is based largely on E. Aust, de ædibus sacris populi Romani unde a primis liberæ reipublicæ temporibus usque ad Augusti imperatoris ætatem Romæ conditis. Marburg, 1889.

[31] Cp. Carter, "Religion of Numa," p. 53 ff.; Wissowa, "Religion und Kultus," p. 198 ff. and in P. W. sub verbo. Diana came into the worship of the league as the goddess of Aricia.

[32] For the question of the exact location of this temple, cp. Jordan-Huelsen, "Topographie," I. 3, p. 158 ff. It is found on fragment 3 of the Forma Urbis Romæ.

[33] On the temple of Apollo, cp. Jordan-Huelsen, "Topographie," p. 535 ff.

posedly protected by the city wall. For the worship of Apollo was purely an affair of the Roman state, and hence could well be inside the wall provided it was outside the pomerium. But again under our present supposition we realize that the Aventine also was a suburb and hence, so far as protection was concerned, it would be a matter of indifference whether the temple was on the Aventine or in the Campus Martius.

Turning from the field of religion to that of constitutional development, it has always been difficult to understand why there should have been only four city tribes, named after the four regions, in case the city so soon extended its borders and took in the Aventine. But if the Aventine was added two centuries later it will readily be seen that the force of habit two centuries old caused the number of city tribes to be limited to four even when the city had exceeded the local limits of the four old regions.

But when we turn to the question of the increase in Rome's population and the disposal of it we have our best argument for treating the Aventine as a suburb. The population was increasing rapidly—we see signs of it in the growing number of foreigners both tradespeople and handicraftsmen. By degrees there arose a problem very similar to that of modern Rome, a dearth of houses for the working classes. It was then (456) that a law was passed providing for the plebeians on the Aventine.[34] Had the Aventine been an internal part of the city it is difficult to see why it would not have been occupied long before. But as an extreme measure the expedient of giving the plebeians land in the suburbs might easily have been adopted.

Thus it was that the city began to outgrow its walls, both in the Aventine region and in the Campus Martius. The proof of this outgrowing is given us in the story of the Gallic catastrophe in B. C. 390. For it is only thus that we can understand why the city was no longer capable of defending itself, and why the Gauls captured it without difficulty, the capitolium alone offering a successful resistance. The tradition of the Gallic catastrophe seems to do

[34] On this law, the lex Icilia, cp. Dionys. 10, 31, and Liv. 3, 31, 1.

violence to the truth in at least two respects; first in underestimating the completeness of the Gallic victory; and second with that sublime indifference to contradiction which is so apt to characterize tradition, by overestimating the amount of physical damage which the Gauls did to the city. At a later time it was customary to attribute all the crookedness and lack of plan which characterized the arrangement of the city streets and buildings to the haste with which Rome was rebuilt after it had been destroyed by the Gauls.[35] But this presupposes that the Gauls wrought an amount of destruction which would partake of an industry quite at variance to what we know of their natural indolence. But quite aside from the question of destruction the Gallic catastrophe had brought one lesson home to the Romans, namely, that their city needed a defence. It is not surprising that in the years following the retreat of the Gauls a new wall was built on a new line so as to include the now populated Aventine. To include the suburb at the south of the Campus Martius was impossible because of engineering difficulties.

It is no wonder therefore that a passage in the sixth book of Livy (chapter 32) dealing with the year B. C. 378 speaks of the building of a wall,[36] and that another passage (Book VII., Chapter 20, under the year B. C. 353) speaks of repairs to walls and towers.[37] Rome was beginning her conquest of Italy, and it was necessary that she should herself be protected from hostile forces. This is accordingly the epoch from which dates the so-called Servian Wall.

[35] Cp. the striking passage in Livy (5, 55): antiquata deinde lege promisce urbs ædificari cœpta. Tegula publice præbita est, saxi materiæque cædendæ, unde quisque vellet, ius factum prædibus acceptis eo anno ædificia perfecturos. Festinatio curam exemit vicos dirigendi, dum omisso sui alienique discrimine in vacuo ædificant. Ea est causa, ut veteres cloacæ, primo per publicum ductæ, nunc privata passim subeant tecta, formaque urbis sit occupatæ magis quam divisæ similis. Cp. also the passage in Tacitus (Annal., 15, 38) where he compares the rebuilding of Rome after the Gallic catastrophe with the rebuilding after Nero's fire.

[36] Et tantum abesse spes veteris levandi fenoris, ut tributo novum fenus contraheretur in murum a censoribus locatum saxo quadrato faciundum.

[37] Legionibusque Romam reductis reliquum anni muris turribusque reficiendis consumptum, et ædis Apollinis dedicata est.

With the capture of the city by the Gauls, Rome enters upon her period of inviolability for almost exactly eight hundred years, and the thought suggests itself irresistibly that the reputation for inviolability thus gained may have been a large factor in preserving her inviolate. Even in these early days the city began to be "that so holy spot, the very Rome."

ROME, April 2, 1909.

THE LINEAR RESISTANCE BETWEEN PARALLEL CONDUCTING CYLINDERS IN A MEDIUM OF UNIFORM CONDUCTIVITY.

By A. E. KENNELLY.

(*Read April 24 1909.*)

It is the purpose of this paper to present formulas and tables for the computation of the linear resistances, conductances and capacities between parallel cylindrical conductors, or between a cylindrical conductor and a parallel indefinitely extending conducting plane. As is shown in the appended bibliography, the problem is by no means new; but the mathematical mode of presentation, and the arithmetical tabulation, here offered, are believed to be new. It is hoped that these will be useful to students of electrical engineering. Antihyperbolic functions are the natural vehicles of expression adapted to this problem.

Infinite Conducting Plane and Parallel Cylinder.

Linear Resistance.—Let a uniform conducting cylinder of radius

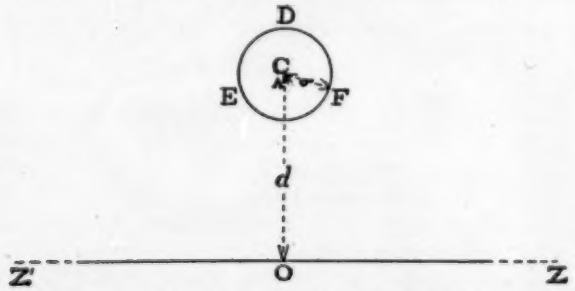


Fig. 1. Section of a conducting cylinder *DEF* parallel to the indefinitely extending conducting plane *Z'OZ*.

σ cm., shown in section at *DEF* in Fig. 1, be situated at an axial distance d cm. from a parallel indefinitely extending conducting

plane $Z'OZ$. Let the space above the plane unoccupied by the cylinder be filled by an indefinitely extending medium of uniform resistivity ρ absohm-cm. Then the linear resistance between the plane and the cylinder, *i. e.*, the resistance of the medium between them, as comprised between a pair of infinite parallel planes perpendicular to the cylinder and 1 cm. apart, will be

$$r_p = \frac{\rho}{2\pi} \cosh^{-1}\left(\frac{d}{\sigma}\right) \quad \begin{matrix}\text{absohm-cms. or C.G.S. magnetic} \\ \text{units of resistance in a linear cm.}\end{matrix} \quad (1)$$

If the conducting surface EDF of the cylinder were unrolled into a flat conducting ribbon $2\pi\sigma$ cm. in breadth, and the ribbon were supported parallel to the plane $Z'OZ$ at a uniform distance $L = \sigma \cosh^{-1}(d/\sigma)$ cm. above it, as indicated in Fig. 2, with vertical insulating side walls, Ez' and Fz, to limit the flow of current through the medium to the parallel distribution shown; then the rectangular slab of medium $EFzz'$ of Fig. 2, would be the equivalent in electric resistance to the indefinitely extending plane and cylinder system of Fig. 1.

In Fig. 2 the depth, or distance across the slab, following the lines of current flow, is $L = \sigma \cosh^{-1}(d/\sigma)$ cm., and the

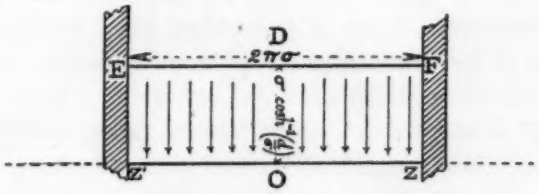


FIG. 2. Equivalent slab section corresponding to infinite plane and parallel cylinder of Fig. 1.

surface area of each face of the slab, per linear cm. of its length, is $S = 2\pi\sigma$ cm.^{2}/cm. so that the linear resistance of the whole is

$$r_p = \frac{L}{S}\rho = \rho \frac{\sigma \cosh^{-1}(d/\sigma)}{2\pi\sigma} = \frac{\rho}{2\pi} \cosh^{-1}\left(\frac{d}{\sigma}\right) \quad \text{absohm-cm.} \quad (2)$$

Since the linear resistance of the plane cylinder system of Fig. 1, or of the slab in Fig. 2, does not depend upon its absolute dimensions, the scale of linear dimensions in the diagram may be chosen

such that $\sigma = 1$ unit, in which case the depth of the slab is $\cosh^{-1} d$ units and the breadth of the slab is 2π units.

The quantity Y defined by the relation

$$Y = \cosh^{-1}(d/\sigma) \qquad \text{numeric (3)}$$

may be called the *distance factor* of the plane-cylinder system; because the distance between electrodes in the equivalent slab of Fig. 2 is

$$L = Y\sigma \qquad \text{cm.}$$

When the radius σ of the cylinder is very small with respect to the distance d; so that d/σ is a large number, we have

$$Y = \log_e \frac{2d}{\sigma} \qquad \text{numeric (4)}$$

so that for such cylinders the linear resistance

$$r_p = \frac{\rho}{2\pi} Y = \frac{\rho}{2\pi} \log_e \frac{2d}{\sigma} \qquad \text{absohm-cm. (5)}$$

The accompanying table gives for successive values of d/σ in column I., the corresponding value of Y in column II. Column III. gives the resistance factor $Y/2\pi$ which, when multiplied by the resistivity ρ of the medium, gives the linear resistance of the plane-cylinder system considered.

Thus, if a conducting cylinder with a radius of 2 cm. is supported at an axial distance of 10 cm. from an infinite conducting plane, in a medium of resistivity $\rho = 3 \times 10^{10}$ absohm-cms., we have $d/\sigma = 5$. The table gives for this ratio the value of Y as 2.2924, and the value of the resistance factor $Y/2\pi = 0.3649$; so that the linear resistance of the system will be $3 \times 10^{10} \times 0.3649 = 1.0947 \times 10^{10}$ absohm-cms.; or 10.947 ohms in a linear cm.

Linear Conductance.—The linear conductance, or conductance per linear cm. of the plane-cylinder system will be by (1)

$$g_p = \frac{2\pi}{\rho \cosh^{-1}(d/\sigma)} = \frac{2\pi}{\rho Y} = \gamma \cdot \frac{2\pi}{Y} \qquad \text{abmhos per cm. (6)}$$

where γ is the uniform conductivity of the medium in abmhos per

cm. The quantity $2\pi/Y$ may be called the *conductance-factor* of the plane-cylinder system. It appears in column V. of the table.

Thus, if a conducting cylinder of radius $\sigma = 0.5$ cm. be supported at an axial distance of $d = 7.5$ cm. from an infinite conducting plane, in a medium of conductivity $\gamma = 10^{-10}$ abmhos per cm., the ratio d/σ in column I. is 15, and the conductance factor for this ratio appears in column V. as 1.848. The linear conductance of the system is thus 1.848×10^{-10} abmhos per cm. The distance-factor of the system is given in column II. as 3.4001; so that the depth of the equivalent rectangular slab of medium is 1.700 cm., the breadth being 3.142 cm.

Linear Electrostatic Capacity.—The linear capacity c_p of a plane-cylinder system in a dielectric medium of specific inductive capacity κ, is numerically the same as the linear conductance of the same system in a medium of conductivity $\kappa/4\pi$ or resistivity $4\pi/\kappa$; so that, in C.G.S. electrostatic units:

$$c_p = \frac{\kappa}{2\cosh^{-1}(d/\sigma)} = \kappa \cdot \frac{1}{2Y} \quad \text{statfarads per cm.} \quad (7)$$

The values of the capacity factor $1/(2Y)$ appear in column VI. of the table for each selected value of d/σ.

Thus, a cylinder of radius $\sigma = 0.4$ cm. is supported at an axial distance of 1 cm. from an infinite conducting plane in a medium of $\kappa = 1$. Here $d/\sigma = 2.5$, and $1/(2Y) = 0.3192$. The linear capacity of the system is therefore 0.3192 statfarad per cm.

In order to convert the linear capacity c_p statfarads per cm. into microfarads per km., expressed by c_p', we have:

$$c_p' = \frac{c_p}{9} = \frac{\kappa}{9} \cdot \frac{1}{2Y} \quad \text{microfarads per km.} \quad (8)$$

Similarly, to express the linear capacity in microfarads per mile

$$c_p'' = \frac{c_p}{5.591} = \frac{\kappa}{5.591} \times \frac{1}{2Y} \quad \text{microfarads per mile} \quad (9)$$

That is, we must divide the capacity-factor of the table by 9 to obtain microfarads per km. or by 5.591 to obtain microfarads per mile.

PROC. AMER. PHIL. SOC., XLVIII. 192 K, PRINTED SEPTEMBER 2, 1909.

Potential Distribution.

On the Median Line Beneath the Cylinder.—It is well known that the flow of electric current, and the distribution of potential, between the conducting cylinder and the plane, are such as might be produced by removing the conducting cylinder and substituting a conducting polar line at A, parallel to the plane. The point A lies on the line OC, and at a distance a from the plane defined by the relation

$$a = \sigma \sinh Y = \sqrt{d^2 - \sigma^2}. \qquad \text{cm. (10)}$$

The values of the polar ratio a/σ are given in the table in column VII. for each of the selected ratios d/σ up to $d/\sigma = 50$, beyond which the difference between a/σ and d/σ is less than 1 part in 5,000. For most practical purposes, it is, therefore, sufficient to regard the polar line as coinciding with the cylinder axis when the distance of that axis from the plane exceeds 50 radii.

In the steady state of flow, the potential at any point y_1 on the line OA (Fig. 3) distant y_1 cm. from O, will be

$$u_1 = I\frac{\rho}{\pi}\tanh^{-1}\left(\frac{y_1}{a}\right) \qquad \text{abvolts (11)}$$

where I is the current strength per linear cm. of the system in absamperes, the potential of the plane $Z'OZ$ being taken as numerically zero.

Similarly, the potential at any other point y_2 on the median line OY, below A, distant y_2 cm. from O, will be:

$$u_2 = I\frac{\rho}{\pi}\tanh^{-1}\left(\frac{y_2}{a}\right) \qquad \text{abvolts (12)}$$

Consequently, if the potential of the surface of the cylinder be u_1, and y_1 be the distance of the lowest point of the cylinder from the plane, the potential of any other point on the line OA between the cylinder and the plane, distant y_2 cm. from the latter, will be:

$$u_2 = u_1\frac{\tanh^{-1}(y_2/a)}{\tanh^{-1}(y_1/a)} \qquad \text{abvolts (13)}$$

Potentials on the Median Line Above the Cylinder.—In the steady state of flow, the potential at any point y_3 on the median line OY, and distant y_3 cm. from O, above the polar point A, is:

$$u_3 = I\frac{\rho}{\pi}\coth^{-1}\left(\frac{y_3}{a}\right) \qquad \text{abvolts (14)}$$

where I and π have the same meanings as above, and the potential of the plane $Z'OZ$ is reckoned as zero.

Similarly, the potential at any other point y_4 on the median line OY, distant y_4 cm. from O, and above the polar point A, is:

$$u_4 = \frac{I\rho}{\pi}\coth^{-}\left(\frac{y_4}{a}\right) \qquad \text{abvolts (15)}$$

Consequently, if the potential of the surface of the cylinder be u_3, and y_3 be the distance of the highest point of the cylinder from the plane, the potential at any other point on the median line, above the cylinder, and distant y_4 cm. from the plane, will be:

$$u_4 = u_3\frac{\coth^{-1}(y_4/a)}{\coth^{-1}(y_3/a)} \qquad \text{abvolts (16)}$$

Potentials at Points Outside the Cylinder and off the Median Line.—If the point in the plane $Z'YZ$ at which the potential is required, lies off the median line OY, the potential may be expressed either:

(a) In terms of rectangular coördinates z and y of the point.

(b) In terms of the ratio of radii vectores to the point, from the polar point A, and from its image.

(a) *Potential in Terms of Rectangular Coördinates.*—Let P, Fig. 3, be the point whose potential is required, and whose rectangular coördinates are y and z, measured respectively along the median line OY, and the line OZ in the infinite conducting plane. Then u, the potential of P, is:

$$u = \frac{I\rho}{2\pi}\tanh^{-1}\left(\frac{2ay}{a^2+y^2+z^2}\right) \qquad \text{abvolts (17)}$$

where I, ρ and a have the values previously assigned, and the potential of the plane $Z'OZ$ is reckoned as zero. Eliminating $I\rho/\pi$ with the aid of (11), we have:

$$u = u_1\frac{\tanh^{-1}\left(\frac{2ay}{a^2+y^2+z^2}\right)}{2\tanh^{-1}(y_1/a)} \qquad \text{abvolts (18)}$$

u_1 is the potential of the conducting cylinder, upon the lowest point of which $y = y_1$ and $z = 0$. Thus, taking the point P in Fig. 3, defined by the coördinates $y = 1$ and $z = 2$, and referring the

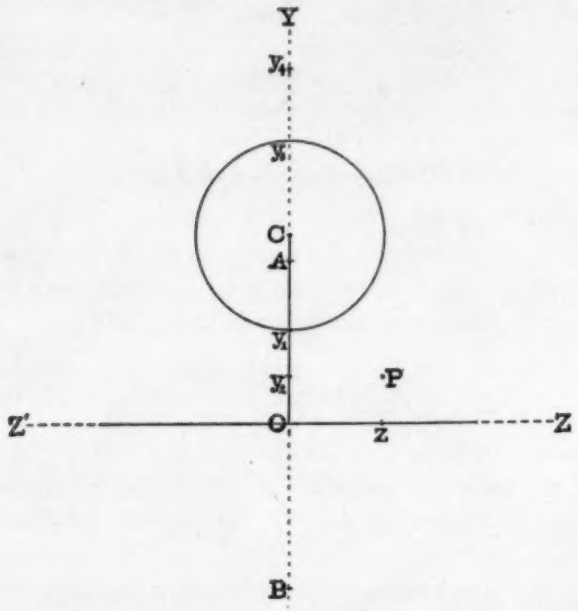


FIG. 3. Coördinates of a point at which the potential is required.

potential u of P to u_1, the potential of the surface of the cylinder, where $y_1 = 2$, $z = 0$, we have $a = 3.4642$ and

$$u = u_1 \frac{\tanh^{-1}(6.9284/17)}{2 \tanh^{-1}(2/3.4642)} = 0.3285 u_1.$$

Formula (18) may also be presented in the form:

$$u = u_1 \frac{\tanh^{-1}\left(\frac{2ay}{a^2 + y^2 + z^2}\right)}{\tanh^{-1}\left(\frac{2ay_1}{a^2 + y_1^2}\right)} \quad \text{abvolts (19)}$$

(*b*) *Potential in Terms of Radii Vectores.*—A line parallel to the axis of the conducting cylinder, drawn through the point B, Fig. 3, on the median line OY and with the distance $OB = OA$, may

be called the *image* of the polar line through OA. The point B, thus defined, may be called the *image polar point*. The points A and B, taken together, may be called the *polar points* of the diagram with respect to the infinite plane and cylinder.

Let P be any point in the plane of the diagram (Fig. 3). Then let r' and r be the lengths of a pair of radii vectores BP, AP, drawn from the polar points B, A, to P respectively. Let these distances $r'r$ be called the *polar distances* of the point P. Then the ratio m of these polar distances will be:

$$m = r'/r \qquad \text{numeric (20)}$$

This ratio may be called the *polar ratio*, for purposes of reference. The polar ratio will manifestly be a number greater than unity for all points in the diagram above the infinite conducting plane $Z'OZ$. It is a well known result that

$$u = \frac{I\rho}{2\pi} \log_e m \qquad \text{abvolts (21)}$$

If a point be selected on the surface of the cylinder, having a potential u_1 abvolts, and for convenience the lowest point of coördinates y_1 and $z = 0$, the polar distances of this point may be denoted by r_1' and r_1; while their ratio may be denoted by $m_1 = r_1'/r_1$. Consequently

$$u_1 = \frac{I\rho}{2\pi} \log_e m_1 \qquad \text{abvolts (22)}$$

and eliminating I, ρ and 2π between (21) (22), we have

$$u = u_1 \frac{\log_e m}{\log_e m_1} = u_1 \frac{\log_{10} m}{\log_{10} m_1} \qquad \text{abvolts (23)}$$

The potential of the infinite plane is here reckoned as zero. It may be observed that

$$m_1 = \frac{r_1'}{r_1} = \frac{a + d - \sigma}{r_1} = \frac{a + d}{\sigma} \qquad \text{numeric (24)}$$

When the cylinder radius is very small, compared with the axial distance d, $d = a$, and

$$m_1 = \frac{r_1'}{r_1} = \frac{2d}{\sigma} = \frac{D}{\sigma} \qquad \text{numeric (25)}$$

It follows from the preceding equations that the equipotential surfaces in an infinite plane-cylinder system are all cylinders having their axes situated on the median line. If u_1 be the potential of the conducting cylinder, and if we denote by Y_1 the value of the distance factor Y for this cylinder, according to formula (3), or to column II. of the table, then the distance factor Y of any cylindrical equipotential surface whose potential is u becomes

$$Y = Y_1 \frac{u}{u_1} \qquad \text{numeric (26)}$$

We have for any such cylinder the equations of condition:

$$Y = \cosh^{-1}(d/\sigma) = \sinh^{-1}(a/\sigma) = \tanh^{-1}(a/d) = \coth^{-1}(d/a) = 2\tanh^{-1}(y/a) \qquad \text{numeric (27)}$$

whence d, the axial distance, or y coördinate, of the cylinder whose potential is u, will be along the median line OY:

$$d = \frac{a}{\tanh\left(Y_1 \frac{u}{u_1}\right)} \qquad \text{cm. (27)}$$

and the radius σ of this equipotential cylinder is:

$$\sigma = \frac{a}{\sinh\left(Y_1 \frac{u}{u_1}\right)} \qquad \text{cm. (28)}$$

The coördinate y of the lowest point of any such equipotential cylinder will be:

$$y = a\left(\frac{m-1}{m+1}\right) \qquad \text{cm. (29)}$$

$$= a\tanh\left(\frac{Y}{2}\right) = a\tanh\left(\frac{Y_1}{2}\frac{u}{u_1}\right) \qquad \text{cm. (30)}$$

so that

$$u = u_1 \frac{\tanh^{-1}\left(\frac{m-1}{m+1}\right)}{\tanh^{-1}(y_1/a)} \qquad \text{abvolts (31)}$$

an expression for the potential of a point in the medium in terms of its polar ratio m, and the distance y_1 of the conducting cylinder from the plane.

The current density δ at any point whose polar distances are r and r' will be perpendicular to the equipotential cylinder passing through the point and will be equal to

$$\delta = I \frac{\rho}{\pi} \cdot \frac{a}{rr'} \quad \text{absamperes per cm.}^2 \quad (31a)$$

The preceding formulas for potential distribution have been developed with reference to a conducting medium between the infinite plane and cylinder. They are, however, applicable to the case of a dielectric medium, if the electric flux ϕ replace the electric current I, and the dielectric constant κ be substituted for γ or $1/\rho$. No substitution will be needed in formulas (13), (16), (18), (19) and (23) to (31), inclusive, which apply either to an insulating or to a conducting medium.

Two Equal and Parallel Conducting Cylinders.

If, instead of an infinite conducting plane and a parallel conducting cylinder, as in Figs. 1 and 3, we have two indefinitely long parallel conducting cylinders of equal diameter, as in Fig. 4, at an interaxial distance CC' of D cm., then each cylinder may be regarded as forming an independent plane-cylinder system with a fictitious infinite conducting midplane $Z'OZ$, axially distant $d = D/2$ cm. from each. This midplane will be perpendicular to the central line CC'. The double-cylinder system will have two polar lines equidistant from the system center O, and represented in Fig. 4 by the polar points AA'. The potential of the midplane $Z'OZ$ will be midway between the potentials of the two cylinders; so that if these have equal and opposite potentials, the potential of the midplane will be zero. All of the preceding formulas for plane-cylinder systems may, therefore, be applied, in duplicate, to the double-cylinder system of Fig. 4.

Linear Resistance of Double Cylinder Systems.—The linear resistance from either cylinder to the midplane is given in formula

(1). Consequently, the linear resistance of the double cylinder system of Fig. 4 is

$$r_{00} = \frac{\rho}{\pi} \cosh^{-1}(d/\sigma) = \frac{\rho}{\pi} Y \qquad \text{absohm-cms.} \quad (32)$$

where $d = D/2$. The resistance factor of the system is thus Y/π, or double that given in column III. of the table.

Thus, if the two cylinders, each of radius $\sigma = 2$ cm. separated

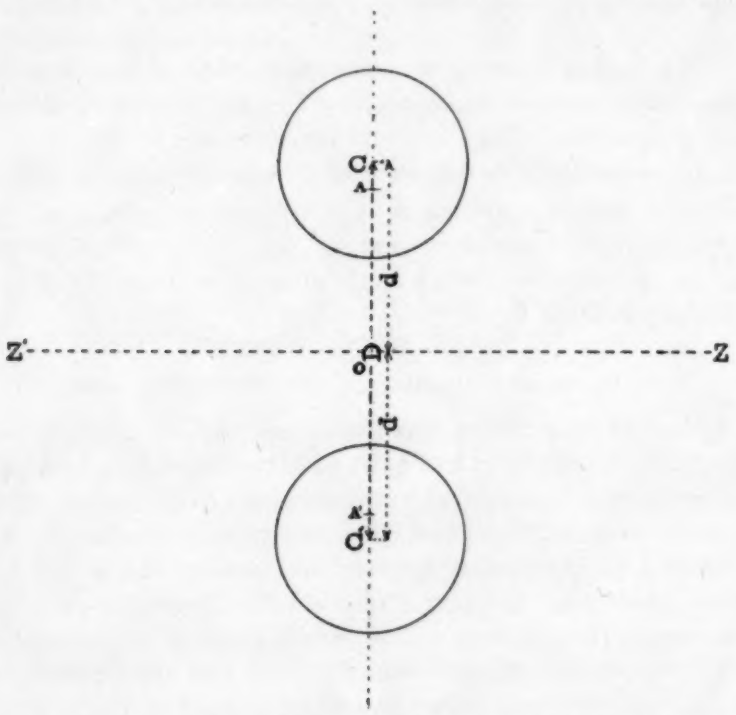


FIG. 4. Two equal and parallel conducting cylinders at interaxial distance of D cm.

by an interaxial distance $D = 8$ cm. in a medium of resistivity $\rho = 5 \times 10^{10}$ absohm-cms. we have $d = 4$, and $d/\sigma = 2$.

$Y = \cosh^{-1} 2 = 1.317$, and the linear resistance is

$$r_{00} = \frac{5 \times 10^{10}}{3.1416} \times 1.317 = 2.096 \times 10^{10} \qquad \text{absohm-cms.}$$

Linear Conductance of Double-Cylinder Systems.—The linear conductance of a double cylinder system will be half that of a plane-cylinder system of equal d/σ; so that:

$$g_{00} = \frac{\pi}{\rho \cosh^{-1}(d/\sigma)} = \frac{\pi}{\rho Y} = \frac{\gamma\pi}{Y} \quad \text{abmhos per cm.} \quad (33)$$

where γ is the conductivity of the medium. The conductance-factor of the double-cylinder system is therefore half of that given in column V. of the table.

Linear Electrostatic Capacity of Double-Cylinder Systems.—The linear capacity C_{00} of a double-cylinder system in a dielectric medium of specific capacity κ is half the capacity of a plane-cylinder system of equal d/σ; so that:

$$c_{00} = \frac{\kappa}{4 \cosh^{-1}(d/\sigma)} = \kappa \cdot \frac{1}{4Y} \quad \text{statfarads per loop cm.} \quad (34)$$

The linear capacity of each cylinder to the zero-potential plane, or the capacity of the system per cylinder-cm., is given by formula (7). The capacity factors of a double-cylinder system of given d/σ are thus half of the values given in column VI. of the table; but the capacity factors of the system per "wire" cm. to zero potential midplane are those recorded in column VI.

At interaxial distances large with respect to the cylinder-radii, $Y = \log_e D/\sigma$, and we obtain the well known formula

$$c_{00} = \frac{\kappa}{4 \log_e (D/\sigma)} \quad \text{statfarads per cm.} \quad (35)$$

The linear capacity of a double-cylinder system expressed in microfarads per km. is

$$c_{00}' = \frac{c_{00}}{9} = \frac{\kappa}{9} \cdot \frac{1}{4Y} \quad \text{microfarads per cm.} \quad (36)$$

Similarly,

$$c_{00}'' = \frac{c_{00}}{5.591} = \frac{\kappa}{5.591} \times \frac{1}{4Y} \quad \text{microfarads per mile} \quad (37)$$

Potential Distribution in Double Cylinder System.—All of the formulas (10) to (31) inclusive referring to the potential distribution in a plane-cylinder system apply immediately to a double-

cylinder system, after the latter has been analyzed into two associated plane-cylinder systems.

Two Unequal Parallel Conducting Cylinders.

Let two parallel conducting cylinders, with their axes at C_1C_2, Fig. 5, have unequal radii σ_1 and σ_2 cm., and be separated by an interaxial distance D cm. If the radii were equal, the midplane $z'z$ would be the plane of zero potential, when the potentials of the cylinders are equal and opposite. The zero-potential plane is, however, displaced from the larger towards the smaller cylinder through a distance of $\Sigma\Delta/2D$ cm.; so that:

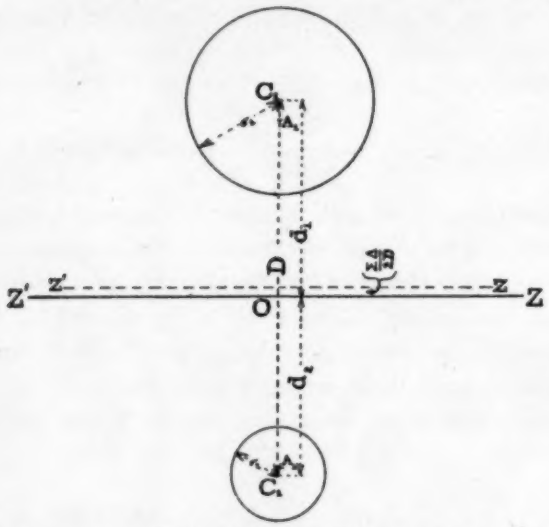


FIG. 5. Two unequal parallel conducting cylinders at interaxial distance of D cm. showing the displacement of the zero-potential plane.

$$\left.\begin{aligned} d_1 &= \frac{D}{2} + \frac{\Sigma\Delta}{2D} \quad \text{cm.} \\ d_2 &= \frac{D}{2} - \frac{\Sigma\Delta}{2D} \quad \text{cm.} \end{aligned}\right\} \tag{38}$$

where $\Sigma = \sigma_1 + \sigma_2$ is the sum and $\Delta = \sigma_1 - \sigma_2$ is the difference of the cylinder radii.

After having established the position of the zero-potential plane $Z'OZ$, the linear resistance between the cylinders may be found by using formula (1) on each side of the plane and adding the two parts. The linear conductance will then be the reciprocal of this result.

The linear capacity of each cylinder to zero-potential plane is to be found by formula (7). The linear capacity per loop cm. may be found from the linear resistance per loop cm. by the formula:

$$c_{00} = \frac{\kappa}{2(Y_1 + Y_2)} \quad \text{statfarads per cm.} \quad (39)$$

For example, if two conducting cylinders of radii $\sigma_1 = 2$ and $\sigma_1 = 1$ cm., respectively, are separated in air by an interaxial distance of 8 cm., the zero-potential plane is displaced through a distance of $\frac{8}{18}$ cm., so that $d_1 = 4\frac{8}{18}$, $d_2 = 3\frac{10}{18}$ cm. The ratio d_1/σ_1 is thus 2.094, and d_2/σ_2 is 3.815. The distance factor Y_1 is 1.37, and Y_2 is 2.014. The linear capacity of C_1 is 0.365 statfarads per cm. and of C_2 0.248 statfarads per cm., each to zero-potential plane. The linear capacity of the pair by (39) is 0.1477 statfarad per loop cm.

The potential distribution in the unequal cylinder system may be obtained as easily as when the cylinders are equal, since the polar points A_1A_2, Fig. 4, lie at equal distances from the zero-potential plane $Z'OZ$.

Excentric Cylinders.

Let the two parallel very thin conducting cylinders be hollow, with radii σ_1 and σ_2. Let one be placed excentrically within the other, as shown in Fig. 6, at an interaxial distance D. Let the line C_1C_2 joining their centers be prolonged as indicated in the figure. The infinite zero-potential plane will perpendicularly intersect this line at an inferred distance of $\Sigma\Delta/2D$ cm. from the middle point of D; so that:

$$d_1 = \frac{\Sigma\Delta}{2D} + \frac{D}{2} \quad \text{cm.} \quad (40)$$

and

$$d_2 = \frac{\Sigma\Delta}{2D} - \frac{D}{2} \quad \text{cm.} \quad (41)$$

The linear resistance between the cylinders can now be determined by finding the linear resistance of each to the infinite conducting plane by formula (1) and then taking the difference between these linear resistances.

Thus, let $\sigma_1 = 4$ cm., $\sigma_2 = 2$ cm., $D = 1$ cm. Then $\Sigma = 6, \Delta = 2$, and $d_1 = 6.5$ cm., $d_2 = 5.5$ cm.

The resistance factor for d_2 by the table is 0.2657.

The resistance factor for d_1 by the table is 0.1697.

The resistance factor between d_2 and d_1 0.0960.

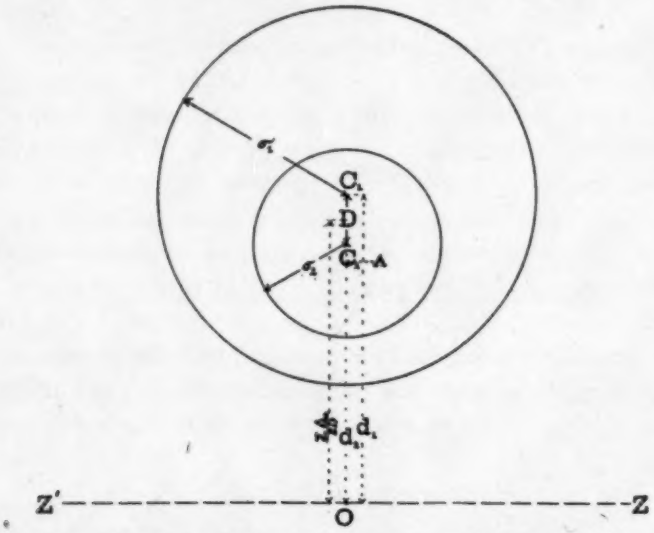


FIG. 6. Two parallel excentric cylinders, one enclosing the other, and the inferred common zero-potential plane.

which multiplied by the resistivity of the medium gives the linear resistance between the cylinders.

Through the use of formulas (40) and (41) all cases of excentric cylinders may be computed by reduction to the equivalent pair of plane-cylinder systems.

GRAPHICAL CONSTRUCTION OF EQUIPOTENTIAL AND STREAM LINES IN A PLANE-CYLINDER SYSTEM.

To draw the equipotential and stream lines of a plane-cylinder system, when the polar distance OA or distance a of the polar axis

from the parallel plane is known, draw zOK, Fig. 7, to represent the plane and on the median line OY, perpendicular to zOK mark off, to scale, the polar distance $a = OA$. Then to locate any equipotential circle of radius $\sigma = OE'$, mark off with center O, a distance $d = OC = AE'$. With center C and the required radius σ, describe the equipotential circle FEB. The distance factor Y for this circle will be expressed by

$$Y = 2 \tanh^{-1}\left(\frac{y_1}{a}\right) \qquad \text{numeric (42)}$$

where y_1 is the distance OF or the y coördinate of the lowest point

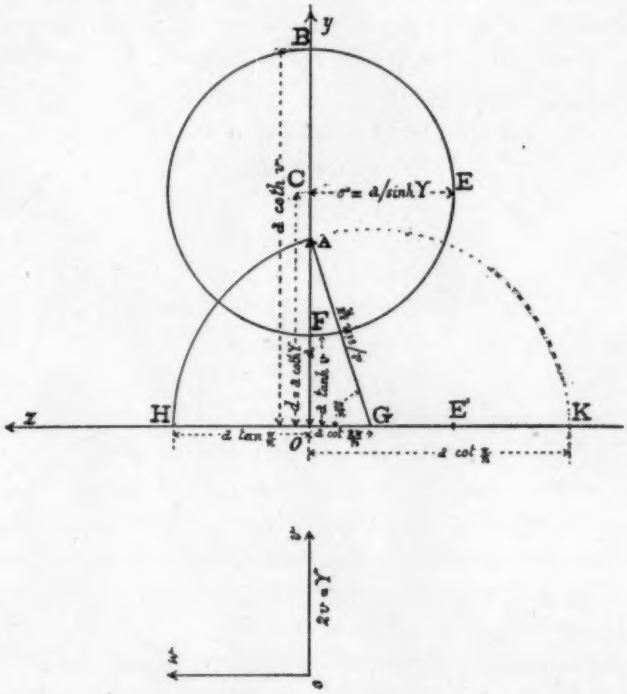


FIG. 7. Diagram for graphic construction of equipotential and stream lines.

on the circle. The potential of the circle with reference to the plane will be

$$u = \frac{I\rho}{2\pi} Y \qquad \text{abvolts (43)}$$

To draw a stream line which shall include with the median line OA the nth part of all the linear flux in the system, mark off on OK a distance $OG = a \cot 2\pi/n$; so that the angle OGA will contain $2\pi/n$ radians. Then with center G and radius GA, describe the circular arc AH, which is the required stream-line.

It may be observed that if we draw two coördinate axes ov ow in the vw plane, the function $\tanh (v + w\sqrt{-1})$ will correspond on the yz plane to the required loci, magnified by a. The locus of this function, when v is given successive constant values and w alone varies, is a series of equipotential circles, while when w is successively assigned constant values and v alone varies, the loci of successive stream-lines are produced. If w is expressed in terms of π as π/n and $2v = Y$, we have

$$OF = a \tanh v = d - \sigma \qquad \text{cm. (44)}$$
$$OB = a \coth v = d + \sigma \qquad \text{cm. (45)}$$
$$CE = a/\sinh Y = \sigma \qquad \text{cm. (46)}$$
$$OC = a \coth Y = d \qquad \text{cm. (47)}$$

also

$$OH = a \tan \pi/n \qquad \text{cm. (48)}$$
$$OK = a \cot \pi/n \qquad \text{cm. (49)}$$
$$GA = a/\sin (2\pi/n) \qquad \text{cm. (50)}$$
$$OG = a \cot 2\pi/n \qquad \text{cm. (51)}$$

Fig. 8 presents the graphical construction of the function $\tanh (v + w\sqrt{-1})$ carried from the vw plane to the yz plane, over the limits $v = -1$ to $v = +1$ and $w = -\pi/2$ to $w = +\pi/2$. The points marked on the vw plane have their corresponding points marked on the yz plane. Thus the point p defined by $v = 1.0$, $w = \pi/2$ on the vw plane is represented by the point p defined by $y = 1.313$, $z = 0$, on the yz plane, or $\tanh (1 + \pi/2 \cdot \sqrt{-1}) = 1.313$. Corresponding areas on the two planes are shaded alike. It follows from the formulas already discussed that linear resistances, conductances and capacities are the same between corresponding conducting surfaces in the two diagrams. Thus, the linear resistance of the double-cylinder system *pqrs—tuvx* is equal to the linear resistance of the rectangular slab system with *pqrs* as one electrode and *tuvx*

as the other; *i. e.*, $2/\pi$ absohm-cm. Moreover, the linear resistance of any curvilinear element, such as between qr on one cylinder, and uv on the other, in the yz system, is equal to the linear resistance between the parallel electrodes qr and uv on the rectilinear vw system ($10/\pi$ absohm-cms. with unit resistivity).

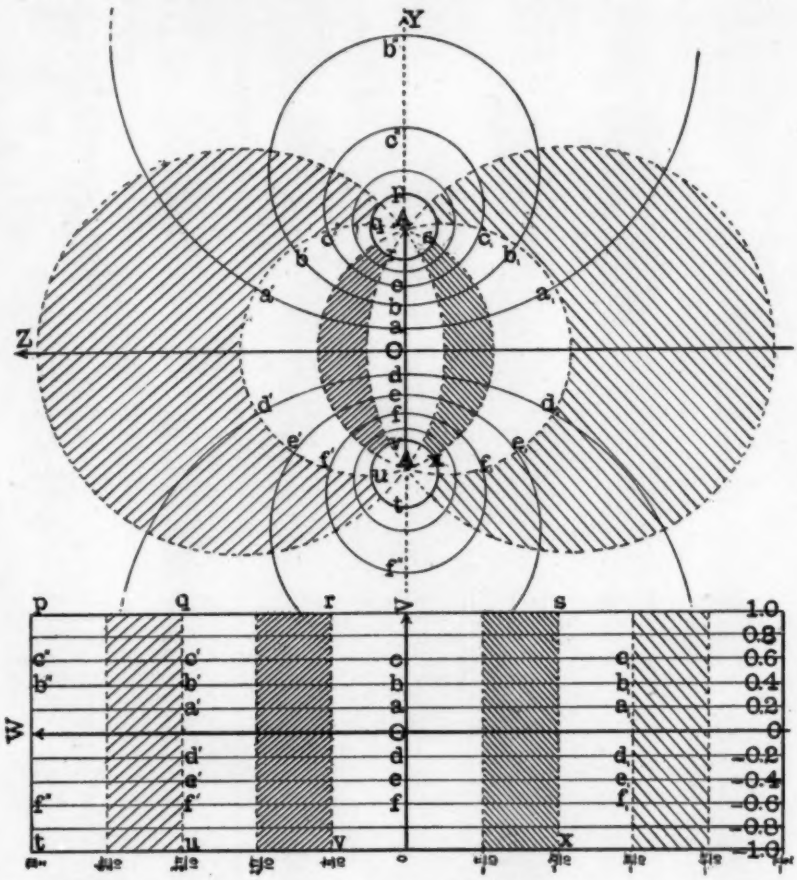


FIG. 8. Graphical comparison of $(v+w\sqrt{-1})$ and of tanh $(v+w\sqrt{-1})$.

In Fig. 8, $a=OA=1$; but it is easy to see that the proposition of equal linear resistances, conductances and capacities between corresponding conductors in the double-cylinder and corresponding rectangular slab systems, is independent of the magnification in the diagram.

I	II Distance Factor $Y=$	III Resistance Factor	IV	V Conductance Factor	VI Capacity Factor	VII sinh Y
d/σ	$\cosh^{-1}\left(\frac{d}{\sigma}\right)$	$Y/2\pi$	$1/Y$	$2\pi/Y$	$1/(2Y)$	$\frac{a}{\sigma}=\sqrt{\left(\frac{d}{\sigma}\right)^2-1}$
1.01	0.1413	0.0225	7.0787	44.47	3.5393	0.1418
1.05	0.3149	0.0501	3.1756	19.95	1.5878	0.3202
1.1	0.4435	0.0706	2.2548	14.16	1.1274	0.4582
1.2	0.6224	0.0991	1.6067	10.095	0.8034	0.6633
1.3	0.7564	0.1204	1.3221	8.307	0.6611	0.8307
1.4	0.8670	0.1380	1.1534	7.246	0.5767	0.9798
1.5	0.9622	0.1531	1.0393	6.531	0.5197	1.1180
1.6	1.0470	0.1666	0.9551	6.002	0.4776	1.2490
1.7	1.1232	0.1788	0.8901	5.594	0.4451	1.3748
1.8	1.1929	0.1899	0.8383	5.267	0.4191	1.4967
1.9	1.2569	0.2001	0.7956	4.999	0.3978	1.6156
2.0	1.3170	0.2096	0.7593	4.771	0.3797	1.7321
2.1	1.3729	0.2185	0.7284	4.576	0.3642	1.8466
2.2	1.4255	0.2296	0.7015	4.407	0.3508	1.9596
2.3	1.4750	0.2348	0.6780	4.259	0.3390	2.0712
2.4	1.5216	0.2422	0.6572	4.129	0.3286	2.1817
2.5	1.5668	0.2494	0.6383	4.010	0.3192	2.2913
2.6	1.6096	0.2562	0.6214	3.903	0.3107	2.4000
2.7	1.6502	0.2626	0.6059	3.807	0.3030	2.5080
2.8	1.6886	0.2688	0.5922	3.721	0.2961	2.6153
2.9	1.7267	0.2748	0.5791	3.639	0.2896	2.7221
3.0	1.7627	0.2806	0.5673	3.564	0.2837	2.8284
3.1	1.7975	0.2861	0.5563	3.495	0.2782	2.9343
3.2	1.8309	0.2914	0.5462	3.432	0.2731	3.0397
3.3	1.8633	0.2966	0.5367	3.372	0.2684	3.1448
3.4	1.8946	0.3015	0.5278	3.317	0.2639	3.2496
3.5	1.9248	0.3063	0.5195	3.264	0.2598	3.3541
3.6	1.9542	0.3110	0.5117	3.215	0.2559	3.4583
3.7	1.9827	0.3156	0.5044	3.169	0.2522	3.5623
3.8	2.0104	0.3200	0.4974	3.126	0.2487	3.6661
3.9	2.0373	0.3242	0.4909	3.084	0.2454	3.7696
4.0	2.0634	0.3284	0.4846	3.045	0.2423	3.8730
4.1	2.0889	0.3325	0.4787	3.008	0.2394	3.9762
4.2	2.1137	0.3364	0.4731	2.973	0.2366	4.0792
4.3	2.1380	0.3402	0.4677	2.939	0.2339	4.1821
4.4	2.1616	0.3440	0.4626	2.907	0.2313	4.2849
4.5	2.1846	0.3477	0.4577	2.876	0.2289	4.3875
4.6	2.2072	0.3513	0.4531	2.847	0.2265	4.4900
4.7	2.2292	0.3548	0.4486	2.819	0.2243	4.5924
4.8	2.2507	0.3582	0.4443	2.792	0.2221	4.6947
4.9	2.2718	0.3616	0.4402	2.766	0.2201	4.7969
5.0	2.2924	0.3649	0.4362	2.741	0.2181	4.8990
5.1	2.3126	0.3681	0.4324	2.717	0.2162	5.0010
5.2	2.3324	0.3712	0.4287	2.694	0.2144	5.1029
5.3	2.3514	0.3743	0.4253	2.672	0.2127	5.2048

I	II Distance Factor $V=\cosh^{-1}\left(\frac{d}{\sigma}\right)$	III Resistance Factor	IV	V Conductance Factor	VI Capacity Factor	VII $\sinh V$
d/σ	V	$V/2\pi$	$1/V$	$2\pi/V$	$1/(2V)$	$\frac{d}{\sigma}=\sqrt{\left(\frac{d}{\sigma}\right)^2-1}$
5.4	2.3709	0.3773	0.4218	2.650	0.2109	5.3066
5.5	2.3895	0.3803	0.4185	2.630	0.2093	5.4083
5.6	2.4078	0.3832	0.4153	2.610	0.2077	5.5100
5.7	2.4258	0.3861	0.4122	2.590	0.2061	5.6116
5.8	2.4435	0.3889	0.4093	2.571	0.2047	5.7131
5.9	2.4608	0.3917	0.4064	2.553	0.2032	5.8146
6.0	2.4779	0.3944	0.4036	2.536	0.2018	5.9161
6.5	2.5590	0.4073	0.3908	2.455	0.1954	6.4226
7.0	2.6339	0.4192	0.3797	2.386	0.1898	6.9282
7.5	2.7036	0.4303	0.3699	2.324	0.1849	7.4330
8.0	2.7687	0.4407	0.3612	2.270	0.1806	7.9373
8.5	2.8297	0.4503	0.3539	2.224	0.1770	8.4410
9.0	2.8873	0.4596	0.3463	2.176	0.1732	8.9443
9.5	2.9417	0.4682	0.3399	2.136	0.1700	9.4472
10.0	2.9932	0.4764	0.3341	2.099	0.1670	9.9499
11	3.0890	0.4916	0.3237	2.034	0.1619	10.9545
12	3.1763	0.5055	0.3148	1.978	0.1574	11.9583
13	3.2566	0.5183	0.3071	1.930	0.1536	12.9615
14	3.3309	0.5301	0.3002	1.887	0.1501	13.964
15	3.4001	0.5411	0.2941	1.848	0.1471	14.967
16	3.4648	0.5514	0.2886	1.814	0.1443	15.969
17	3.5255	0.5611	0.2837	1.782	0.1418	16.971
18	3.5827	0.5702	0.2791	1.754	0.1396	17.972
19	3.6369	0.5788	0.2750	1.728	0.1375	18.974
20	3.6882	0.5870	0.2712	1.704	0.1356	19.975
21	3.7371	0.5948	0.2676	1.681	0.1338	20.976
22	3.7837	0.6022	0.2643	1.661	0.1321	21.977
23	3.8282	0.6093	0.2612	1.641	0.1306	22.978
24	3.8708	0.6161	0.2584	1.623	0.1292	23.979
25	3.9116	0.6226	0.2557	1.606	0.1278	24.980
26	3.9509	0.6287	0.2531	1.590	0.1266	25.981
27	3.9887	0.6348	0.2507	1.575	0.1254	26.981
28	4.0250	0.6406	0.2485	1.561	0.1243	27.982
29	4.0604	0.6462	0.2463	1.548	0.1232	28.983
30	4.0941	0.6516	0.2443	1.535	0.1221	29.983
32	4.1590	0.6619	0.2404	1.511	0.1202	31.984
34	4.2193	0.6715	0.2370	1.489	0.1185	33.985
36	4.2765	0.6806	0.2338	1.469	0.1169	35.986
38	4.3306	0.6892	0.2309	1.451	0.1155	37.987
40	4.3819	0.6972	0.2282	1.434	0.1141	39.987
42	4.4307	0.7051	0.2257	1.418	0.1129	41.988
44	4.4772	0.7126	0.2234	1.403	0.1117	43.989
46	4.5217	0.7196	0.2212	1.390	0.1106	45.989
48	4.5642	0.7264	0.2191	1.377	0.1096	47.990
50	4.6051	0.7329	0.2172	1.364	0.1086	49.990

PROC. AMER. PHIL. SOC., XLVIII. 192 I, PRINTED SEPTEMBER 3, 1909.

I d/σ	II Distance Factor $Y = \cosh^{-1}\left(\frac{d}{\sigma}\right)$	III Resistance Factor $Y/2\pi$	IV $1/Y$	V Conductance Factor $2\pi/Y$	VI Capacity Factor $1/(2Y)$	VII $\sinh Y$ $\frac{d}{\sigma} = \sqrt{\left(\frac{d}{\sigma}\right)^2 - 1}$
52	4.6443	0.7392	0.2153	1.353	0.1077	52
54	4.6821	0.7452	0.2136	1.342	0.1068	54
56	4.7184	0.7509	0.2119	1.332	0.1060	56
58	4.7535	0.7565	0.2104	1.322	0.1052	58
60	4.7874	0.7619	0.2088	1.312	0.1044	60
65	4.8676	0.7747	0.2054	1.291	0.1027	65
70	4.9416	0.7864	0.2024	1.272	0.1012	70
75	5.0106	0.7975	0.1996	1.254	0.0998	75
80	5.0751	0.8077	0.1970	1.238	0.0985	80
85	5.1358	0.8173	0.1947	1.224	0.0974	85
90	5.1930	0.8264	0.1926	1.210	0.0963	90
95	5.2470	0.8350	0.1906	1.198	0.0953	95
100	5.2983	0.8433	0.18874	1.1859	0.09437	100
110	5.3936	0.8585	0.18540	1.1648	0.09270	110
120	5.4806	0.8723	0.18246	1.1464	0.09123	120
130	5.5607	0.8852	0.17983	1.1298	0.08992	130
140	5.6348	0.8969	0.17747	1.1150	0.08874	140
150	5.7038	0.9078	0.17532	1.1016	0.08766	150
160	5.7683	0.9180	0.17336	1.0892	0.08668	160
170	5.8290	0.9278	0.17156	1.0778	0.08578	170
180	5.8861	0.9369	0.16989	1.0674	0.08495	180
190	5.9402	0.9456	0.16834	1.0577	0.08417	190
200	5.9915	0.9536	0.16690	1.0486	0.08345	200
220	6.0868	0.9688	0.16429	1.0322	0.08215	220
240	6.1738	0.9827	0.16197	1.0176	0.08099	240
260	6.2538	0.9954	0.15990	1.0047	0.07995	260
280	6.3279	1.0071	0.15803	0.9930	0.07902	280
300	6.3969	1.0180	0.15633	0.9822	0.07817	300
320	6.4615	1.0283	0.15476	0.9725	0.07738	320
340	6.5221	1.0381	0.15322	0.9634	0.07666	340
360	6.5793	1.0471	0.15199	0.9550	0.07600	360
380	6.6333	1.0557	0.15075	0.9473	0.07538	380
400	6.6846	1.0639	0.14960	0.9400	0.07480	400
420	6.7334	1.0716	0.14851	0.9332	0.07426	420
440	6.7799	1.0790	0.14749	0.9268	0.07375	440
460	6.8244	1.0862	0.14653	0.9207	0.07327	460
480	6.8669	1.0929	0.14563	0.9151	0.07282	480
500	6.9078	1.0993	0.14476	0.9096	0.07238	500
550	7.0031	1.1146	0.14279	0.8972	0.07140	550
600	7.0901	1.1284	0.14104	0.8862	0.07052	600
650	7.1701	1.1411	0.13947	0.8764	0.06974	650
700	7.2442	1.1530	0.13804	0.8674	0.06902	700
750	7.3132	1.1640	0.13674	0.8591	0.06837	750
800	7.3778	1.1741	0.13554	0.8518	0.06777	800
850	7.4384	1.1838	0.13444	0.8449	0.06722	850

I d/σ	II Distance Factor $Y = \cosh^{-1}\left(\frac{d}{\sigma}\right)$	III Resistance Factor $Y/2\pi$	IV $1/Y$	V Conductance Factor $2\pi/Y$	VI Capacity Factor $1/(2Y)$	VII $\sinh Y$ $\frac{a}{\sigma} = \sqrt{\left(\frac{d}{\sigma}\right)^2 - 1}$
900	7.4955	1.1930	0.13341	0.8383	0.06671	900
950	7.5496	1.2016	0.13246	0.8323	0.06623	950
1000	7.6009	1.2097	0.13156	0.8266	0.06578	1000
1100	7.6962	1.2249	0.12993	0.8165	0.06497	1100
1200	7.7832	1.2387	0.12848	0.8074	0.06424	1200
1300	7.8633	1.2515	0.12717	0.7990	0.06359	1300
1400	7.9374	1.2632	0.12599	0.7916	0.06300	1400
1500	8.0064	1.2742	0.12490	0.7848	0.06245	1500
1600	8.0709	1.2845	0.12390	0.7786	0.06195	1600
1700	8.1315	1.2940	0.12298	0.7728	0.06149	1700
1800	8.1887	1.3032	0.12212	0.7674	0.06106	1800
1900	8.2428	1.3118	0.12132	0.7624	0.06066	1900
2000	8.2941	1.3200	0.12056	0.7575	0.06028	2000
2100	8.3428	1.3278	0.11986	0.7532	0.05993	2100
2200	8.3894	1.3351	0.11920	0.7490	0.05960	2200
2300	8.4338	1.3423	0.11857	0.7451	0.05929	2300
2400	8.4764	1.3490	0.11798	0.7414	0.05899	2400
2500	8.5172	1.3555	0.11741	0.7378	0.05871	2500
2600	8.5564	1.3618	0.11687	0.7344	0.05844	2600
2700	8.5942	1.3678	0.11636	0.7312	0.05818	2700
2800	8.6305	1.3735	0.11587	0.7280	0.05794	2800
2900	8.6656	1.3791	0.11540	0.7251	0.05770	2900
3000	8.6995	1.3845	0.11495	0.7224	0.05748	3000
3100	8.7323	1.3898	0.11452	0.7196	0.05726	3100
3200	8.7641	1.3949	0.11410	0.7170	0.05705	3200
3300	8.7948	1.3996	0.11370	0.7144	0.05685	3300
3400	8.8247	1.4045	0.11332	0.7121	0.05666	3400
3500	8.8537	1.4090	0.11295	0.7098	0.05648	3500
3600	8.8818	1.4135	0.11259	0.7075	0.05630	3600
3700	8.9092	1.4180	0.11224	0.7053	0.05612	3700
3800	8.9359	1.4220	0.11191	0.7032	0.05596	3800
3900	8.9619	1.4262	0.11158	0.7012	0.05579	3900
4000	8.9872	1.4302	0.11127	0.6992	0.05564	4000
4100	9.0118	1.4342	0.11097	0.6973	0.05549	4100
4200	9.0360	1.4381	0.11067	0.6954	0.05534	4200
4300	9.0595	1.4419	0.11038	0.6936	0.05519	4300
4400	9.0825	1.4456	0.11010	0.6918	0.05505	4400
4500	9.1050	1.4491	0.10983	0.6902	0.05492	4500
4600	9.1270	1.4526	0.10957	0.6885	0.05479	4600
4700	9.1485	1.4560	0.10931	0.6869	0.05466	4700
4800	9.1695	1.4593	0.10906	0.6853	0.05453	4800
4900	9.1901	1.4627	0.10881	0.6838	0.05441	4900
5000	9.2103	1.4659	0.10857	0.6822	0.05429	5000

Notation.

a = polar distance or distance of polar axis from parallel plane in a plane-cylinder system, cm.
c_p = linear capacity of plane-cylinder system, statfarads/cm.
c_p' = linear capacity of plane-cylinder system, microfarads/km.
c_p'' = linear capacity of plane-cylinder system, microfarads/mile
c_{00} = linear capacity of double-cylinder system, statfarads/cm.
c_{00}' = linear capacity of double-cylinder system, microfarads/km.
c_{00}'' = linear capacity of double-cylinder system, microfarads/mile
d = distance of cylinder axis from plane, cm.
$d_1 d_2$ = distances of cylinder axes from plane in double-cylinder system with unequal cylinders, cm.
D = $2d$ or interaxial distance between two cylinders in a double cylinder system, cm.
$\Delta = \sigma_1 - \sigma_2$ = difference in radii of two cylinders, cm.
δ = current density at a point in the medium, absamperes/cm.^{2}.
g_p = linear conductance of plane-cylinder system, abmho/cm.
g_{00} = linear conductance of double-cylinder system, abmho/cm.
κ = specific inductive capacity of medium,
γ = conductivity of medium, abmho/cm.
I = linear current in a system, absamperes/cm.
L = length of flux paths in rectangular slab, cm.
$m = r'/r$, polar ratio, or ratio of vector lengths from poles to a point in the medium, numeric
$1/n$ = a fractional part of the total linear flux, limited by a stream line.
π = 3.14159 ....
$r_1 r'$ = polar distances or vector lengths from poles to a point.
r_p = linear resistance of a plane-cylinder system absohm/cm.
r_{00} = linear resistance of a double-cylinder system, absohm/cm.
ϕ = linear electric flux in a system, statmaxwells/cm.
ρ = resistivity of medium, absohm-cm.
S = linear surface area of a conducting slab, cm.^{2}/cm.
$\Sigma = \sigma_1 + \sigma_2$ = sum of radii of two unequal cylinders, cm.
σ = radius of a cylinder, cm.
u = potential of a cylinder, abvolts or statvolts
vw = rectangular coördinates of points in a plane, cm.
Y = distance factor of a system = $\cosh^{-1}(d/\sigma)$, numeric
yz = rectangular coördinates of points in a plane, cm.
$y_1 y_2$ = y-coördinates of points on median line below a cylinder, cm.
$y_3 y_4$ = y-coördinates of points on median line above a cylinder, cm.

Bibliography.

Kirchhoff, Dr. S.

1845. Über den Durchgang eines elektrischen Stromes durch eine Ebene insbesondere durch eine kreisformige. *Poggendorf's Annalen*, 1845, Vol. 44, pp. 497–514.

Smaasen, Dr. W.

1846. Vom dynamischen Gleichgewicht der Electricität in einer Ebene oder einem Körper. *Poggendorf's Annalen*, 1846, Vol. 69, pp. 161–180.

Vom dynamischen Gleichgewicht der Elektricität in einem Körper und in unbegränzten Raum.

Ridolfi.

1847. *Il Cimento*, An V., 1847, May–June.

Kirchhoff.

1870. Carl's *Repertorium für experim. Physik*, Vol. 6, 1870, p. 11.

Gaugain.

1862. *Ann. de Chim. et de Physique*, 1862, Ser. 3, Vol. 66, p. 203.

Blavier.

Resistance Electrique de l'Espace compris entre deux cylindres, quoted by Gaugain in 1862; also *Journal de Physique*, Vol. 3, p. 115, April, 1874.

Smith, W. Robertson.

1869–70. *Proc. Edin. Roy. Soc.*, 1869–70, pp. 79–99.

Foster, G. C., and Lodge, O.

1875. On the Flow of Electricity in a uniform plane Conducting Surface. *Phil. Mag.*, 1875, 4th Ser., Vol. 49, pp. 385–400 and 453–471.

Heaviside, O.

1880. The Electrostatic Capacity of Suspended Wires. *Jour. Soc. Tel. Engrs.*, 1880, Vol. 9, p. 115. *Electrical Papers.* London, Vol. I, pp. 42–46.

Kennelly, A. E.

1892. The Problem of Eccentric Cylinders. *The Electrical World*, N. Y., 1892, Vol. 20, pp. 338–339.

Houston, E. J., and Kennelly, A. E.

1894. The Inductance and Capacity of Suspended Wires. *The Electrical World*, 1894, Vol. 24, No. 1, p. 6.

Lichtenstein, Leo.

1904. Über die rechnerische Bestimmung der Capacität von Luftleitern und Kabeln. E. T. Z., 1904, p. 126.

Benischke.

1907. Die Wissenschaftlichen Grundlagen der Elektrotechnik, Berlin, 1907, p. 44–46.

ON AN ADJUSTMENT FOR THE PLANE GRATING SIMILAR TO ROWLAND'S METHOD FOR THE CONCAVE GRATING.[1]

By CARL BARUS.

(*Read April 24, 1909.*)

1. *Apparatus.*—The remarkable refinement which has been attained (notably by Mr. Ives and others) in the construction of celluloid replicas of the plane grating, makes it desirable to construct a simple apparatus whereby the spectrum may be shown and the measurement of wave-length made, in a way that does justice to the astonishing performance of the grating. We have, therefore, thought it not superfluous to devise the following inexpensive contrivance, in which the wave-length is strictly proportional to the shift of the carriage at the eye-piece; which for the case of a good 2-meter scale divided into centimeters, admits of a measurement of wave-length to a few Ångström units and with a millimeter scale should go much further.

Observations are throughout made on both sides of the incident rays and from the mean result most of the usual errors should be eliminated by symmetry.

In Fig. 1, *A* and *B* are two double slides, like a lathe bed, 155 cm. long and 11 cm. apart, which happened to be available for optical purposes, in the Laboratory. They were therefore used, although single slides at right angles to each other, similar to Rowland's, would have been preferable. The carriages *C* and *D*, 30 cm. long, kept at a fixed distance apart by the rod *aRb*, are in practice a length of ¼-inch gas pipe, swivelled at *a* and *b*, 169.4 centimeters apart, and capable of sliding right and left and to and fro, normally to each other.

[1] The investigations in this paper were undertaken throughout in conjunction with my son, Mr. Maxwell Barus; but it seemed advisable that I should undertake the publication in these PROCEEDINGS myself, with the present acknowledgment.

The swivelling joint which functioned excellently, is made very simply of ¼-inch gas pipe T's and nipples, as shown in Fig. 2. The lower nipple *N* is screwed tight into the T, but all but tight into the carriage *D*, so that the rod *ab* turns in the screw *N*, kept oiled. Similarly the nipple *N''* is either screwed tight into the T (in one

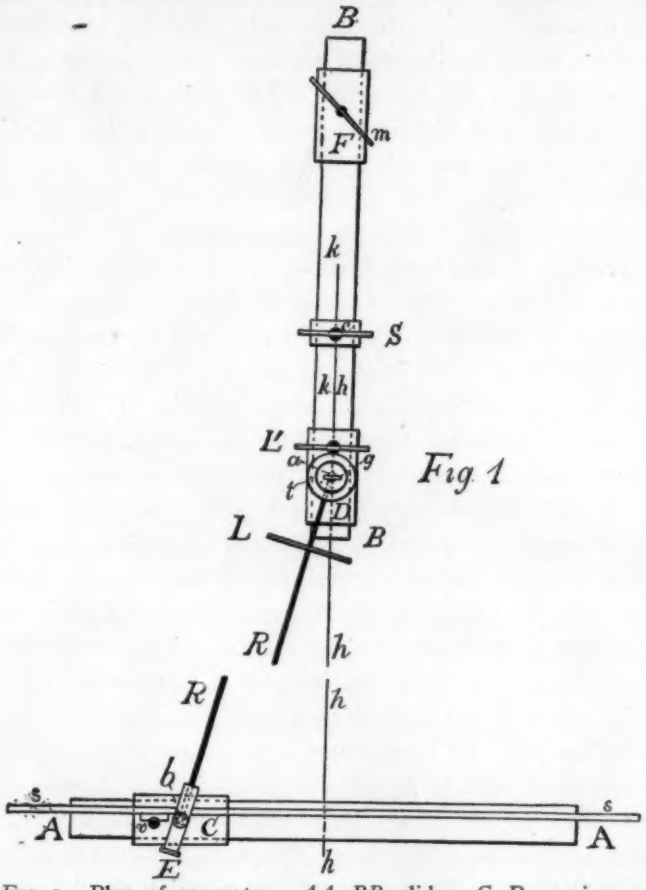


Fig. 1

FIG. 1. Plan of apparatus. *AA*, *BB*, slides; *C*, *D*, carriages; *R*, connecting rod.

method, revoluble grating), or all but tight (in another method, stationary grating), so that the table *tt*, which carries the grating *g* may be fixed while the nipple *N''* swivels in the T. Any ordinary

laboratory clamp *K* and a similar one on the upright *c* (screwed into the carriage *S*) secures a small rod *k* for this purpose. Again a hole may be drilled through the standards at *K* and *c* and provided with set screws to fix a horizontal rod *k* or check. The rod *k* should be long enough to similarly fix the standard on the slide *S* carrying the slit and be prolonged further toward the rear to carry the flame or Geissler tube apparatus. The table *tt* is revoluble on a brass rod fitting within the gas pipe, which has been slotted across so that the conical nut *M* may hold it firmly. The axis passes through the middle of the grating, which is fastened centrally to the table *tt* with the usual tripod adjustment.

2. *Single Focusing Lens in Front of Grating.*—I shall describe three methods in succession, beginning with the first. Here a large lens *L*, of about 56 cm. focal distance and about 10 cm. in diameter, is placed just in front of the grating, properly screened and throwing an image of the slit *S* upon the cross-hairs of the eye-piece *E*, the line of sight of which is always parallel to the rod *ab*, the end *b* swivelled in the carriage *C*, as stated (see Fig. 2). An ordinary lens of 5 to 10 cm. focal distance, with an appropriate diaphragm, is adequate and in many ways preferable to stronger eye-pieces. The slit *S*, carried on its own slide and capable of being clamped to *c* when necessary, as stated, is additionally provided with a long rod *hh* lying underneath the carriage, so that the slit *S* may be put accurately in focus by the observer at *C*. *F* is a carriage for the mirror or the flame or other source of light whose spectrum is to be examined; or the source may be adjustable on the rear of the rod by which *D* and *S* are locked together.

Finally the slide *AB* is provided with a scale *ss* and the position of the carriage *C* read off by aid of the vernier *v*. A good wooden scale graduated in centimeters happened to be available, the vernier reading to within one millimeter. For more accurate work a brass scale in millimeters with an appropriate vernier should of course be used.

Eye-piece *E*, slit *S*, flame *F*, etc., may be raised and lowered by the split tube devise shown as at *M* and *M'* in Fig. 2.

3. *Adjustments.*—The first general test which places slit, grating and its spectra and the two positions of the eye-piece in one plane,

is preferably made with a narrow beam of sunlight, though lamplight suffices in the dark. Thereafter let the slit be focused with the eye-piece on the right marking the position of the slit; next focus the slit for the eye-piece on the left; then place the slit midway between these positions and now focus by slowly rotating the grating. The slit will then be found in focus for both positions

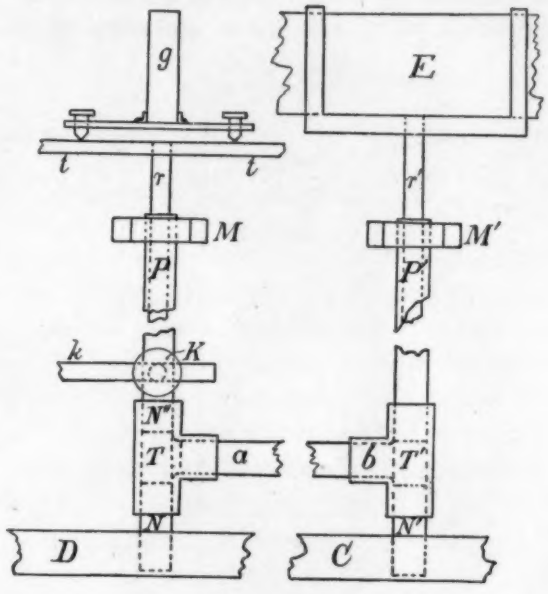


Fig. 2.

FIG. 2. Elevation of the grating (g) and the eyepiece (E) standards.

and the grating which acts as a concave lens counteracting L will be symmetrical with respect to both positions.

Let the grating be thus adjusted when fixed normally to the slide B or parallel to A. Then for the first order of the spectra the wave-length $\lambda = d \sin \theta$, where d is the grating space and θ the angle of diffraction. The angle of incidence i is zero.

Again let the grating, adjusted for symmetry, be free to rotate with the rod ab. Then θ is zero and $\lambda = d \sin i$.

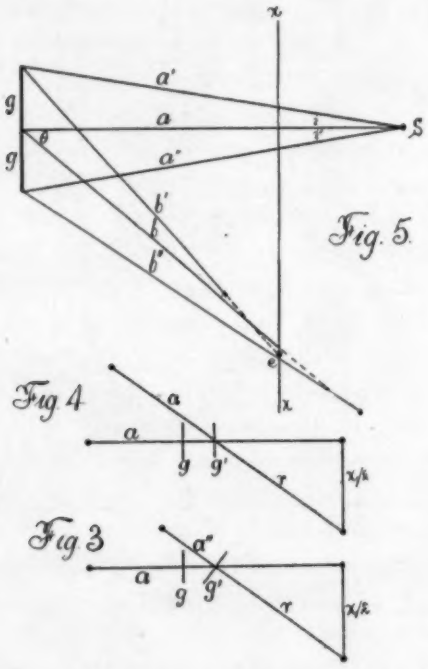
In both cases however if $2x$ be the distance apart of the car-

riage C, measured on the scale ss, for the effective length of rod $ab = r$ between axis and axis,

$$\lambda = dx/r \quad \text{or} \quad (d/2r)2x,$$

so that in either case λ and x are proportional quantities.

The whole spectrum is not however clearly in focus at one time, though the focusing by aid of the rod hh is not difficult. For extreme positions a pulley adjustment, operating on the ends of h



FIGS. 3, 4, 5. Diagrams.

is a convenience, the cords running around the slide AA. In fact if the slit is in focus when the eye-piece is at the center ($\theta = 0$, $i = 0$) at a distance a from the grating, then for the fixed grating, Fig. 4,

$$a' = a \frac{r^2}{r^2 - x^2},$$

where a' is the distance between grating and slit for the diffraction corresponding to x. Hence the focal distance of the grating regarded as a concave lens is $f' = ar^2/x^2$. For the fixed grating and a given color, it frequently happens that the undeviated ray and the diffracted rays of the same color are simultaneously in focus, though this does not follow from the equation.

Again for the rotating grating, Fig. 3, if a'' is the distance between slit and grating

$$a'' = a\frac{r^2 - x^2}{r^2},$$

so that its focal distance is

$$f'' = a\frac{r^2 - x^2}{x^2},$$

It follows also that $a' \times a'' = a^2$. For $a = 80$ cm. and sodium light, the adjustment showed roughly $f' = 650$ cm., $f'' = 570$, the behavior being that of weak concave lenses. The same $a = 80$ cm. and sodium light showed furthermore $a' = 91$ and $a'' = 70.3$.

Finally there is a correction needed for the lateral shift of rays, due to the fact that the grating film is enclosed between two moderately thick plates of glass (total thickness $t = .99$ cm.) of the index of refraction n. This shift thus amounts to

$$e = \frac{tx}{r}\left(\frac{1}{\sqrt{1 - x^2/r^2}} - \frac{1}{\sqrt{n^2 - x^2/r^2}}\right)\frac{b}{a}.$$

But since this shift is on the rear side of the lens L, its effect on the eye-piece beyond will be (if f is the principal focal distance and b the conjugate focal distance between lens and eye-piece, remembering that the shift must be resolved parallel to the scale ss)

$$e = \frac{tx}{r}\left(\frac{1}{\sqrt{1 - x^2/r^2}} - \frac{1}{\sqrt{n^2 - x^2/r^2}}\right)\left(\frac{b}{f} - 1\right),$$

where the correction e is to be added to $2x$, and is positive for the rotating grating and negative for the stationary grating.

Hence in the mean values of $2x$ for stationary and rotating grating the effect of e is eliminated. For a given lens at a fixed distance from the eye-piece $(b/f - 1)$ is constant.

4. *Data for Single Lens in Front of Grating.*—In conclusion we select a few results taken at random from the notes.

Grating	Line.	Observed $2x'$.	Shift.	Corrected $2x$.
Stationary	C	132.60	—.26	132.34
	D_2	118.90	—.23	118.67
	F	98.23	—.19	98.04
	Hydrogen Violet	87.87	—.16	87.71
Rotating	C	132.10	+.26	132.36
	D_2	118.45	.23	118.68
	F	97.90	.19	98.09
	H. Violet	87.50	.16	87.66

The real test is to be sought in the coresponding values of $2x$ for the stationary and rotating cases, and these are very satisfactory, remembering that a centimeter scale on wood and a vernier reading to millimeters only was used for measurement.

5. *Single Focusing Lens Behind the Grating.*—The lens L', which should be achromatic, is placed in the standard behind g. The light which passes through the grating is now convergent, whereas it was divergent in § 2. Hence the focal points at distances a', a'' lie in front of the grating; but in other respects the conditions are similar but reversed. Apart from signs, for the stationary grating

$$a' = a\frac{r^2 - x^2}{r^2},$$

and for the rotating grating

$$a'' = a\frac{r^2}{r^2 - x^2}.$$

The correction for shift loses the factor $(b/f - 1)$ and becomes

$$e = \frac{tx}{r}\left(\frac{1}{\sqrt{1 - x^2/r^2}} - \frac{1}{\sqrt{n^2 - x^2/r^2}}\right).$$

As intimated, it is negative for the rotating grating and positive for the stationary grating. It is eliminated in the mean values.

6. *Data. Single Lens Behind the Grating.*—An example of the results will suffice. Different parts of the spectrum require focusing.

Grating.	Line.	$2x'$	Shift.	$2x$
Stationary	D_2	118.40	+.13	118.53
Rotating	D_2	118.65	—.13	118.52

The values of $2x$, remembering that a centimeter scale was used, are again surprisingly good. The shift is computed by the above equation. It may be eliminated in the mean of the two methods. The lens L' may be more easily and firmly fixed than L.

7. *Collimator Method.*—The objection to the above single-lens methods is the fact that the whole spectrum is not in sharp focus at once. Their advantage is the simplicity of the means employed. If a lens at L' and at L are used together, the former as a collimator (achromatic) and with a focal distance of about 50 cm., and the latter (focal distance to be large, say 150 cm.) as the objective of a telescope, all the above difficulties disappear and the magnification may be made even excessively large. The whole spectrum is brilliantly in focus at once and the corrections for the shift of lines due to the plates of the grating vanish. Both methods for stationary and rotating gratings give identical results. The adjustments are easy and certain, for with sunlight (or lamplight in the dark) the image of the slit may be reflected back from the plate of the grating on the plane of the slit itself, while at the same time the transmitted image may be equally sharply adjusted on the focal plane of the eye-piece. It is therefore merely necessary to place the plane of spectra horizontal. Clearly a' and a'' are all infinite.

In this method the slide S and D are clamped at the focal distance apart, so that flame, etc., slit, collimator lens and grating move together. The grating may or may not be revoluble with the lens L on the axis a.

8. *Data for the Collimator Method.*—The following data chosen at random may be discussed. The results were obtained at different times and under different conditions. The grating nominally contained about 15,050 lines per inch. The efficient rod length ab was $R = 169.4$ cm. Hence if $1/C = 15{,}050 \times .3937 \times 338.8$, the wavelength $\lambda = C \cdot 2x$ cm.

Grating.	Lines.	$2x'$	$2x$
Stationary	D_2	118.30	118.19
Rotating	D_2	118.08	118.19
Stationary	D_2	118.27	118.16
Rotating	D_2	118.05	118.16

Rowland's value of D_2 is 58.92×10^{-6} cm.; the mean of the two values of $2x$ just stated will give 58.87×10^{-6} cm. The difference may be due either to the assumed grating space, or to the value of R inserted, neither of which were reliable absolutely to much within .1 per cent.

Curious enough an apparent shift effect remains in the values of $2x$ for stationary and rotating grating, as if the collimation were imperfect. The reason for this is not clear, though it must in any case be eliminated in the mean result. Possibly the friction involved in the simultaneous motion of three slides is not negligible and may leave the system under slight strain equivalent to a small lateral shift of the slit.

9. *Discussion.*—The chief discrepancy is the difference of values for $2x$ in the single lens system (for D_2, 118.7 and 118.5 cm., respectively) as compared with a double lens system (for D_2', 118.2 cm.) amounting to .2 to .4 per cent. For any given method this difference is consistently maintained. It does not, therefore, seem to be mere chance.

We have for this reason computed all the data involved for a fixed grating 5 cm. in width, in the two extreme positions, Fig. 5, the ray being normally incident at the left hand and the right hand edge respectively for the method of § 6. The meaning of the symbols is clear from Fig. 5, S being the virtual source, g the grating, e the diffraction conjugate focus of S for normal incidence, so that $b = r$ is the fixed length of rod carrying grating and eye-piece. It is almost sufficient to assume that all diffracted rays b' to b'' are directed towards e, in which case equations (1) would hold; but this will not bring out the divergence in question. They were therefore not used. Hence the following equations (2) to (5) successively apply where d is the grating space.

(1) $\cot\theta' = (b/g + \sin\theta)/\cos\theta$; $\cot\theta'' = (b/g - \sin\theta)/\cos\theta$;

(2) $a = b/\cos^2\theta$; $a' = a'' = \sqrt{g^2 + a^2}$;

(3) $\sin i' = \sin i'' = g/a'$;

(4) $-\sin i' + \sin(\theta + \theta') = \lambda/d$; $\sin\theta = \lambda/d$;
$\sin i'' + \sin(\theta - \theta'') = \lambda/d$;

(5) $\cos^2 i'/a' = \cos^2(\theta + \theta')/b'$; $\cos^2 i''/a'' = \cos^2(\theta - \theta'')/b''$.

Since θ, g, λ, d, b, are given θ' and θ'' are found in equation (4), apart from signs. If δ_1 and δ_1'' be the distance apart of the projections of the extremities of b' and b, b and b'', respectively, on the line x,

$$\begin{aligned} \delta_1' &= g + (b - b') \sin\theta - b' \sin i' \\ \delta_1'' &= g + (b'' - b) \sin\theta - b'' \sin i'' \end{aligned} \tag{6}$$

If δ_2' and δ_2'' be the distance apart of the intersections of the prolongation of b' and b, b and b'', respectively, with the line x,

$$\begin{aligned} \delta_2' &= \sin(\theta + \theta')(b\cos\theta/\cos(\theta + \theta') - b') \\ \delta_2'' &= \sin(\theta - \theta'')(b'' - b\cos\theta/\cos(\theta - \theta'')) \end{aligned} \tag{7}$$

Given $b = 169.4$ cm., $\theta = 20° 22'$, about for sodium, $g = 5$ cm., the following values are obtained:

$$\begin{array}{lll} \theta' = 1° 36', & a = 192.7 \text{ cm.}, & b' = 166.0 \text{ cm.}, \\ \theta'' = 1° 34', & a' = a'' = 192.8 \text{ cm.}, & r = b = 169.4 \text{ cm.}, \\ i' = i'' = 1° 30', & & b'' = 172.4 \text{ cm.}, \end{array}$$

whence

$$\delta_1' = 1.92 \text{ cm.}, \quad \delta_2'' = 1.74 \text{ cm.}$$

These limits are surprisingly wide. If, however, they should be quite wiped out on focusing, for any group of rays and symmetrical observations on the two sides of the apparatus, this would be no source of discrepancy. The effect of focusing the two parts of the grating may, in the first instance, be considered as a prolongation of b' till it cuts x, together with the corresponding points for the intersection of b'' with x. Thus the values δ_2' and δ_2'' are here in question and they are

$$\delta_2' = 1.97 \text{ cm.}, \qquad \delta_2' - \delta_1' = .05 \text{ cm.}$$

whence

$$\delta_2'' = 1.65 \text{ cm.}, \qquad \delta_1'' - \delta_2 = .09 \text{ cm.}$$

are the conjugate foci for the extreme rays of the grating, respectively, beyond the conjugate focus of the middle or normal rays b, on x. Hence the mean of the extreme rays lies at .07 cm. beyond

(greater θ) the normal ray and the λ found in the first instance is too large as compared with the true value for the normal ray.

The datum .07 cm. may be taken as the excess of $2x$, corresponding to the excess of angle for a grating one half as wide and observed on both sides ($2x$), as was actually the case. Finally, since the whole of the grating is not in focus at once a correction less than .07 cm. for $2x$ must clearly be in question. This is quite below the difference of several millimeters brought out in §§ 4 and 6.

To make this point additionally sure and avoid the assumption of the last paragraph, we will compute the conjugate focus of the central ray (different angles θ) on the b' focal plane parallel to the grating and to x and on the b'' focal plane parallel to x. The computation is simpler if the central ray is thus focused, than if the extreme rays are focused on the x plane. The distance apart will be

$$\delta_3' = g - b' \cos(\theta + \theta')(\tan(\theta + \theta') - \tan\theta),$$
$$\delta_3'' = g - b'' \cos(\theta - \theta'')(\tan\theta - \tan(\theta - \theta'')).$$

Inserting the results for θ, θ_1', θ_1'', b', b'', g,

$$\delta_3' = .06, \quad \delta_3'' = -.04.$$

Both the b foci thus correspond to large angles. Their mean, however, may be considered as vanishing on the intermediate x plane.

Thus it is clear that the effect of focusing is without influence on the diffraction angle and much within the limits of observation. It is therefore probable that the residual discrepancy in the three methods is referable to a lateral motion of the slit itself due to insufficient symmetry of the slides AA and BB in the above adjustment. This agrees, moreover, with the residual shift observed in the case of parallel rays in § 8.

Brown University,
Providence, R. I.

THE ELECTRON METHOD OF STANDARDIZING THE CORONAS OF CLOUDY CONDENSATION.

By CARL BARUS.

(*Read April 24, 1909.*)

1. *Introductory.*—Last year I published some preliminary experiments[1] in which the coronal display of the fog chamber was standardized by aid of the value of Thomson's electron, $10^{10}e = 3.4$ electrostatic units, and of the known velocity of the ions. Later similar experiments were made in terms of the former datum and the decay constants of the ions, though this method is not here to be considered. In the experiments in question a separate leaded condenser was used to determine the ionization, while the nucleation was measured in a cylindrical fog chamber. The data, though necessarily rough, owing to the dampness of the room in the summer time, when used for the determination of e by aid of my earlier and independent constants of coronas, nevertheless gave a series of promising values. In the paper cited it was assumed that the whole current due to both positive and negative ions is measured. If, however, the current observed is due to negative ions, while the negative ions only were caught in the fog chamber used, as now appears probable, then the data would be (V denoting the fall of potential per second, dV/dr the average field, all referred to volts, N the number of nuclei (negative ions), per cubic centimeter).

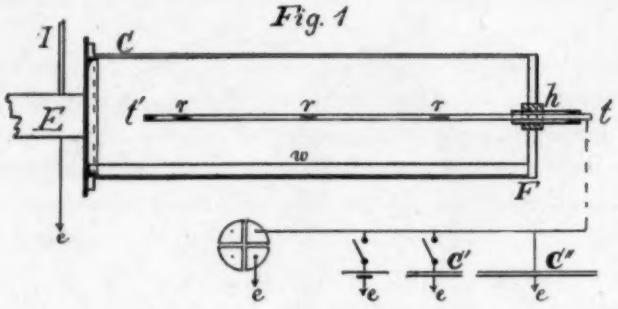
dV/dr	$10^3 \dot{V}/V$	N	$10^{10}e$
1.0	40	150,000	3.3
.7	50	185,000	3.2
.7	60	210,000	3.7
1.2	137	570,000	2.9

where the velocity of negative ions in a unit field of dry air is taken as $v = 1.87$ cm./sec.

[1] *American Journal of Science*, XXVI., 1908, p. 87; *idem*, p. 324.

In the following experiments I have returned to the measurements of N in terms of e and the velocities of the ions, modifying the method by using the cylindrical fog chamber both as an electrical condenser for the measurement of current, as well as for the specification of the number of ions in action by aid of the coronas of cloudy condensation.

Fig. 1



2. *Apparatus.*—This consists of a cylinder of glass C, F, about 45 cm. long, 13.4 cm. internal diameter, closed at one end F and provided with a brass cap C, with exhaust E and influx attachments I, in the usual way. There is a layer of water w at the bottom. The glass must be scrupulously clean within; and this is best secured by scouring with a probang of soft rubber under water, until the water adheres as an even film on shaking. The fog chamber is put to earth, as at e.

The end F is perforated at h, to receive the aluminum tube tt', closed at t' and open at t, 40 cm. long and .64 cm. external diameter. Sealed tubelets of radium $r, r, \ldots$ may be placed at intervals within this tube to ionize the surrounding wet air. The walls being about .1 cm. thick, β and γ rays are wholly in question. Neither emanation nor α rays escaped the double thickness of aluminum. The tube tt' is grasped at t by a sheath of hard rubber with an annular air space and fixed in place by a rubber cork. If care be taken to keep the tube in dry air except when in use, there is no conduction leakage of consequence.

The end t, moreover, is placed in connection with a Dolezalek electrometer, by aid of a thin wire (not shown) running axially

within an earthed tin drain pipe and away from the fog chamber, to escape the action of γ rays as much as possible. In fact their combined effect does not exceed 2 per cent. and is determined in special measurements.

The keys to the electrometer,[2] etc., were all placed on pillars of hard rubber and actuated by long wooden rods from a distance. So far as possible the electrical wires of the room were surrounded by earthed pipes, but it was not practicable to carry this out completely so that a method of correction appears in the work below. Even when the electric lighting circuit was completely cut out, the electrostatic drift in question remained.

The measurements were standardized and the electric system charged by a Carhart-Clarke cell.

The radium tubelets used were as follows:

No. I, 100 milligrams, strength 10,000 ×
No. II, 10 milligrams, strength 200,000 ×
No. III, 100 milligrams, strength 10,000 ×
No. IV, 100 milligrams, strength 7,000 ×
No. V, 100 milligrams, strength 20,000 ×

3. *Electrical Condensers.*—To give the fall of potential a suitably small value relatively to the period of the damped drop of the needle, a number of auxiliary condensers, C', C'', Fig. 1, are needed. It suffices, however, to measure three capacities, viz.,

1. That of the cored fog chamber alone, c;
2. That of a relatively large auxiliary condenser, including the electrometer, the piped wires and the fog chamber, $C'' + c$;
3. That of a standard condenser, C', for reference.

In the present paper C' was computed by the equation

$$C' = \frac{A}{4\pi d}\left(1 + \frac{1}{\sqrt{\pi A}}\left(d + d \ln \frac{16\sqrt{\pi A}(d + d')}{d^2} + d' \ln \frac{d + d'}{d}\right)\right)$$

where A is the area, d the distance apart and d' the thickness of

[2] The disposition of condensers C', C'', cell, etc., earthed at e is suggested in Fig. 1.

the brass plates. Since A is equal 315 sq. cm., $d = .082$ cm., $d' = .67$ cm.,

$$C' = 305.6(1 + .0784) = 330 \text{ cm.}$$

This value will suffice for the present purposes, though it needs further correction by comparison with a standard condenser, not now at hand.

A special key was provided (Fig. 1) whereby C' could be switched into the electrometer system or out of it and put to earth. Hence in a series of successive discharges

$$(C'' + c)V = (C'' + C' + c)V',$$
$$(C'' + c)V' = (C'' + C' + c)V'',$$

etc., so that for n discharges, if the residual potential is V_n

$$V(C'' + c)^n = V_n(C'' + C' + c)^n,$$

from which the total capacity $C = C'' + C' + c$ is determinable in terms of C'. The results were:

Positive charge, $C'' + C' + c = 1{,}445, \quad 1{,}443, \quad 1{,}422,$
Negative charge, $C'' + C + c = \quad 1{,}482, \quad 1{,}480,$
Mean $C = 1{,}459,$

the experiments alternating from positive to negative charge, because of the marked drift by the electrometer system when isolated from the cell, as already specified. To measure the small capacities c, of the fog chamber, the same method with *ten* discharges suffices, if C'' is excluded and C' retained. Thus the data were successively found,

$+$ Charge, $c = 11.8 \quad 12.4 \quad 12.2 \quad 12.9,$
$-$ Charge, $c = 10.8 \quad 10.4 \quad 11.1 \quad 11.5,$
Mean $c = 11.3 \quad 11.4 \quad 11.6 \quad 11.2,$

eliminating the drift in the final mean, $c = 11.4$.

Since the capacity c in terms of the effective internal radius R_2 and external radius R_1 the length l of the clindrical condenser may be written

$$\frac{1}{l} \log \frac{R_1}{R_2} = \frac{1}{2 \times 2.3 \times c},$$

the constant c furnishes a mean value for the factor on the left. The ratio of 4.6 c to the measured value of $(\log R_1/R_2)/l$ was .568, a reduction factor used throughout the tables below.

4. *Method Pursued.*—If C is equal to $C''+C'+c$ we may write the equation for the negative ionization N (positive charge)

$$N=\frac{C\ln R_1/R_2}{600\pi lve}\frac{d(\ln V)}{dt}=\kappa\frac{d(\ln V)}{dt},$$

where R_1, R_2 and l are the effective radii and length of the condenser, $10^{10}e=3.4$, $v=1.51$ cm./sec., and $u=1.37$ cm./sec., the velocity of the negative and positive ions in the unit field, volt/cm., in case of moist air. The factor $(\ln R_1/R_2)/l$ is replaced by $1/2C$, as specified in § 3, which must here be regarded as an adequate correction for the ends and the imperfect cylindricity of the condenser fog chamber.

Similarly the equation for the positive ionization is (negative charge),

$$N'=\frac{C\ln R_1/R_2}{600\pi lue}\frac{d(\ln V')}{dt}=\kappa'\frac{d(\ln V')}{dt},$$

and the total ionization is therefore $N+N'$.

The experiments below will show that even if the fog chamber is put to earth, there is a drift towards negative potential, sufficiently steady to be eliminated in the mean results. Hence if V_0 be the effective negative potential of the wet glass envelope we may write tentatively,

$$N=\kappa\frac{d\ln(V-V_0)}{dt}\quad\text{or}\quad N\left(1-\frac{V_0}{V}\right)=\kappa\frac{d(\ln V)}{dt}$$

where V_0 is intrinsically negative.

Similarly,

$$N'\left(1+\frac{V_0}{V}\right)=\kappa'\frac{d(\ln V')}{dt}.$$

Hence if $V=V'$, $N+N'$ the total ionization is again

$$\frac{d}{dt}(\kappa\ln V+\kappa'\ln V').$$

Direct experiments, however, show that the drift results from

the influx of a high permanent positive voltage. Curiously enough even when the lighting circuit is cut out, the effect remained with undiminished intensity. It will appear elsewhere, that in the absence of radium and of initial charge in the condenser, the equation $I_0 = C\dot{V}_0$ where $\dot{V}_0$ for any given ionization is a constant negative quantity, applies very closely within the limits of measurable $\dot{V}_0$ values. Hence in the presence of radium in the core of the cylindrical fog chamber and a positive charge,

$$I_0 + 600\pi lVNev/(\ln R_1/R_2) = C\dot{V}.$$

Thus in this case

$$NV = \kappa d(V - V_0)/dt; \quad -N'V' = \kappa' d(-V' - V_0)/dt,$$

and for the same $V = V'$, to a first approximation

$$N + N' = d(\kappa \ln V + \kappa' \ln V')_{V=V'}/dt,$$

as before. If the equation for N is integrated and $N/\kappa = K$, since $I_0 = C\dot{V}_0$, $\dot{V}_0$ being intrinsically negative,

$$V = \epsilon^{-Kt}(V_c - \dot{V}_0/K) + \dot{V}_0/K; \quad V' = \epsilon^{-K't}(V_c' + V_0/K') - \dot{V}_0/K',$$

where V_c and V_c' are the initial positive and negative potentials. The constant $\dot{V}_0$ increases with the strength of the ionization but has a fixed value for a given ionization.

5. *Data: High Ionization: Currents*—The tables[3] investigated contain the mean potentials $\bar{V}$, the positive and negative logarithmic currents $d(\log V)/dt$ (apart from the constant), the apparent nucleation N positive and N' negative, computed from these data and additional information as to conduction leakage and effect of γ rays. In most of the cases the corresponding logarithmic currents due to γ rays outside the fog chamber was carefully measured in the same units, by placing a short hard-rubber rod between the end t of the aluminum tube, Fig. 1, and the wire leading to the electrometer. This cuts out the fog chamber but leaves the whole remaining circuit undisturbed. Similarly the leak value of $d(\log V)/dt$ in the absence of radium and due to mere conduction of moist parts is always quite negligible. Thus in the data in

[3] The tables will be removed for brevity, as Figs. 2–4 sufficiently reproduce the data.

question for relative logarithmic currents of the order of .035, the γ ray effect is .0010, the conduction leakage smaller than .0001. The other extreme, *i. e.*, the value of $d(\log V)/dt$ for the freely falling needle is about .1 in the same units. Hence it follows that if the needle falls faster than would be quite trustworthy, the auxiliary capacity selected is too small. The time interval between observations for V was 4 sec., throughout.

6. *The Same: Coronas.*—These results (to be given in Figs. 2*a* and 2) contain the data for the maximum ionizations obtainable with the radium tubelets I., II., III., IV., V. at my disposal. The corresponding corona was a large orange-yellow type, representing (in my former reductions) 506,000 nuclei in the exhausted fog chamber. I have supposed this to be equivalent to 653,000 when the fog chamber is at atmospheric pressure, seeing that the coronas are actually displaced during exhaustion; *i. e.*, at the maximum ionization does not coincide in the position with the largest corona on exhaustion,[4] but is displaced in the direction of the exhaust currents. The observation would seem to mean that exhaustion is more rapid than the reproduction of ions to restock the region of dilatation. In general this inherent discrepancy of a marked distribution of ionization increasing from end to end of the fog chamber is still outstanding. It is partially allowed for since the observations are made near the middle of the chamber where the average conditions supervene.

7. *The Same: Summary.*—The data given in Fig. 2*a* merely show the fall of potential in scale readings, in the successive observations 4 seconds apart, for positive and negative charges. Fig. 2 gives the corresponding positive and negative *apparent ionizations.* If the two curves between .8 and 1.2 volts be considered, the mean ionization of each is

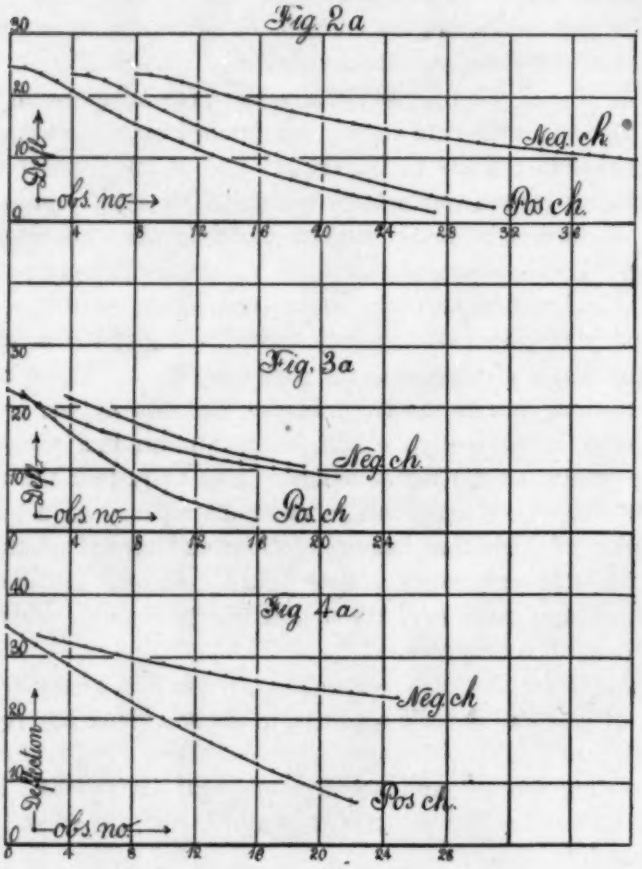
Apparent positive ions (negative charge), $N =$	540,000.
Apparent negative ions (positive charge), $N' =$	1,164,000.
Total true ionization, $N + N' =$	1,704,000.
Total nuclei caught,	650,000.

It will be seen that $N + N'$ is the true total ionization, positive

[4] See papers cited; also *Science,* XXVIII., p. 26, 1908.

and negative, if $10^{10}e = 3.4$. Only 65/170, or about 38 per cent., of this is actually caught in the given fog chamber on exhaustion, provided the old coronal values are corect.

If, however, it is assumed that negative ions only are caught



during exhaustion in the fog chamber in question, then the value of the electron would be

$$10^{10}e = 3.4 \times 2.62 \times \tfrac{1}{2} = 4.4 \text{ electrostatic units.}$$

The irregularities of the curves, Fig. 2, are due in part to fluctuations of the drift and in part to errors inevitable in derivations so

close together; but such errors necessarily compensated each other in the mean values.

10. *Data: Moderate Ionization: Electrical Currents.*—These results were obtained by placing but one radium tubelet, No. IV., in the aluminum tube tt' of the condenser-fog-chamber. The data were found in the same way as in the above. $N = \kappa d(\log V)/dt$, as usual.

Both positive and negative currents were observed in succession and the true total ionization is $N + N'$ as before. Moreover, the capacity of the condensers were widely varied, 410 to 1,459 cm., without showing serious divergences.

11. *The Same: Coronas.*—At a fall of pressure of 21 cm. (and somewhat below) or $\delta p/p = .27$, the nucleation was stationary and equal to $N = 113{,}000$ in the exhausted fog chamber. At atmospheric pressure therefore $113{,}000 \times 1.37 = 154{,}000$ nuclei should have been present. The effect of a charge on the core of the condenser did not appreciably diminish the nucleation.

12. *The Same: Summary.*—The successive observations in scale parts at intervals, 30 seconds apart, are shown in Fig. 3*a*, the slopes only being of interest. The apparent values of N are given in Fig. 3. All the four series show about the same drift, even though taken many days apart. The condenser effect (excessive rapidity of needle) may be considered eliminated for capacities greater than 500 cm.

By averaging the ionizations between $\overline{V} = .6$ and $\overline{V} = 1.24$ in both curves the data found are as follows:

Apparent negative ions, $N =$	278,000.
Apparent positive ions, $N' =$	107,000.
True total ions, $N + N' =$	385,000.
Total nuclei,	180,000.

Hence about 47 per cent. of all the ions were caught on exhaustion, if the values of u, v, e, N, inserted, are correct. Supposing that negative ions only are caught in the above fog chamber, the value of the electron would be

$$10^{10}e = 3.4 \times 2.14 \times \tfrac{1}{2} = 3.6 \text{ electrostatic units.}$$

Fig. 2.

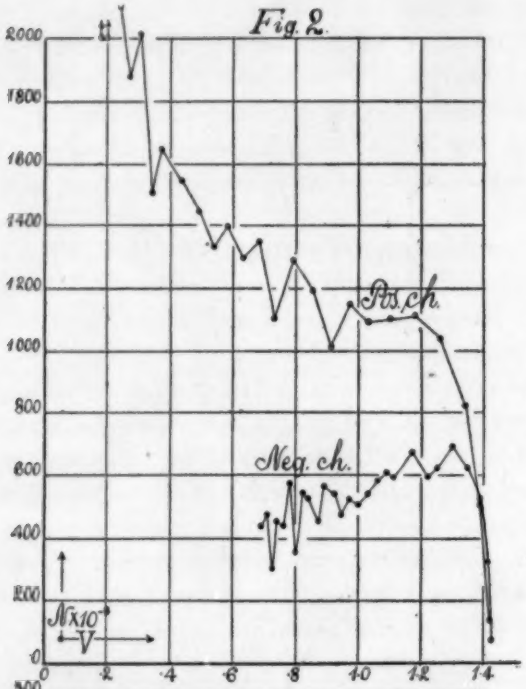


Fig. 4.

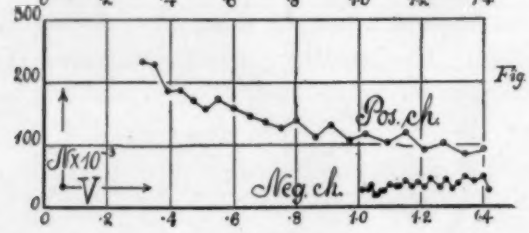
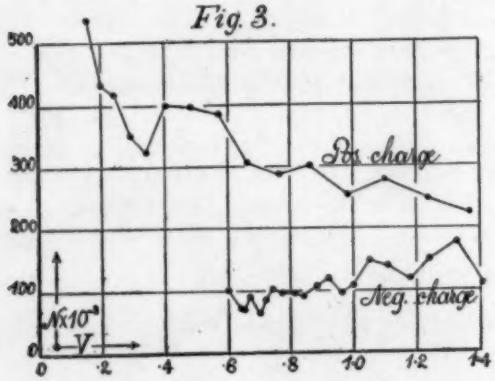


Fig. 3.



13. *Data: Small Ionization: Electric Currents.*—In the next series of experiments the aluminum tube tt', Fig. 1, was surrounded by a lead tube with walls .117 cm. thick, leaving the γ rays only effective and these much reduced in intensity. The data are sufficiently given in the following charts.

14. *The Same: Coronas.*—The coronas found at a drop of pressure similar to the above $\delta p/p = .300$, corresponded in my tables to 46,200 nuclei in the exhausted fog chamber. Hence at atmospheric pressure there should have been 64,000. The effect of charging the core was not definite.

15. *The Same: Summary.*—The drop of potential in scale parts, in successive intervals 30 cm. apart, is given in Fig. 4*a*, showing how much slower the negative charges are lost than the positive charges. The apparent values of N are given in Fig. 4, to which remarks similar to those already made are applicable. There is the usual drift and the usual temporary fluctuation.

If the mean data be taken between $V = 1.1$ and 1.4 volts, the results are

Apparent positive ions, N'	$=$ 37,000.
Apparent negative ions, N	$=$ 98,000.
True total ionization, $N + N'$	$=$ 135,000.
Total nuclei caught,	60,000.

It follows, then, that about 44 per cent. of the total ionization computed from $10^{10}e = 3.4$, u and v, is caught on condensation.

If we suppose the negative ions only are caught in the above fog chamber the electron value is

$$e \times 10^{10} = 3.4 \times 2.3 \times \tfrac{1}{2} = 3.9 \text{ electrostatic units.}$$

Conclusion.—Supposing the electron value to be $10^{10}e = 3.4$ electrostatic units as before, the normal velocities of the ions in wet air to be $u = 1.37$, $v = 1.51$ cm./sec., in the volt/cm. field, the coronal equivalent of the ions caught in the above fog chamber is in the several cases,

Total ions, 1,700,000,	Total nuclei, 38 per cent.
385,000,	47 per cent.
135,000,	44 per cent.

When N is 1,700,000 the coronas are too diffuse for sharp specification. If it is assumed that negative ions only are caught, and if the nucleations corresponding to the coronas seen in the given fog chamber be taken as developed in my earlier work, then for

$$N + N' = 1{,}700{,}000, \quad 385{,}000, \quad 135{,}000,$$

the electron values are

$$10^{10}e = 4.4, \quad 3.6, \quad 3.9,$$

electrostatic units.

With regard to the two parts of this paper that need revision the first, the comparison of the computed condenser capacity C' with a standard, is a minor matter; but the other, *i. e.*, the marked distribution of ionization along the axis of the fog chamber, will need further inquiry. In the direction of the exhaustion the amount of ionization may vary in the ratio of more than 1 to 2, in a fog chamber of about one half meter of length; and this under conditions where there should apparently be no variations and irrespective of the production of radiation from within or from outside of the fog chamber.

Brown University,
Providence, R. I.

THE ELECTROMETRIC MEASUREMENT OF THE VOLTAIC POTENTIAL DIFFERENCE, BETWEEN THE TWO CONDUCTORS OF A CONDENSER, CONTAINING A HIGHLY IONIZED MEDIUM.

By CARL BARUS.

(*Read April 24, 1909.*)

1. *Introductory.*—The difficulties encountered in the preceding paper (§ 4), were made the subject of direct investigation by replacing the fog chamber with a metallic cylindrical condenser, the core of which was an aluminum tube, 50 cm. long and .63 cm. in diameter, the shell a brass tube, 50 cm. long and 2.1 cm. in diameter, coaxial with the former. Sealed radium tubelets could be placed within the aluminum tube, or withdrawn from it. Moreover, either the outer coat or the core of the condenser could be joined in turn with the Dolezalek electrometer, the other being put to earth. The conducting system now appears as follows (Fig. 1), C being the outer coat or brass shell, A the aluminum core and r the radium tubes in the cylindrical core. Conductors are earthed at e. BB show the metallic connections with the auxiliary condensers C', C''. E is one of the insulated quadrants of the electrometer with the highly charged needle N, E being virtually also a condenser.

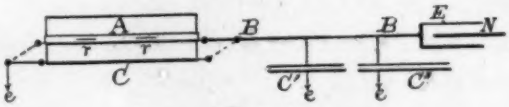


Fig. 1

A Clark standard cell may be inserted for standardization, but it is otherwise withdrawn.

Direct experiment showed the self charging tendencies to come apparantly from the highly charged needle N, as if positive ions were loged into the conductor $EBBA$ for a positive needle, negative ions

for a negative needle. In addition to this however there is a *voltaic* difference, aluminum-brass, at AC when radium is in place and the medium therefore highly ionized. The latter potentials are usually negligible. These are the chief electromotive forces, the first very high (150 volts) and in a weakly ionized medium; the other low (.2 volt) but in an intensely ionized medium: thus they may produce equal currents. Other voltages such as the room potential may be operative, but their effect is secondary. If the capacities C', C'', are successively removed the electrometer current increases proportionately, showing its origin to be directed from the needle toward the insulated or non-earthed pair of quadrants.

If the condenser metals are reversed (see Fig. 1), the voltaic couple is reversed. This makes it possible to obtain both the voltaic contact potential and the ionization in the condenser C, from a pair of commutated measurements.

2. *Theory.*—Let V_n be the potential at the electrometer, V_c the voltaic potential difference of the two metals of the condenser, V the potential of the insulated conductor BB, measured by the electrometer. Let n be the hypothetical ionization in the electrometer, N the (radium) ionization in the condenser (length l, radii R_1, R_2). Let C be the total capacity of the systems $CBBE$. Then

$$\dot{V} = A(V_n - V)n - \frac{600\pi lNev}{C \ln R_1/R_2}(V - V_c)$$

where A is a constant, u and v the normal velocities of the positive and negative ions, e the charge of the electron. The needle is positively charged. This may be written

$$\dot{V} = \dot{V}_a - K(V - V_c),$$

where for $N = 0$, $K = 0$, or

$$\dot{V} = \dot{V}_a = A(V_n - V)n,$$

i. e., the current in the electrometer, observed in the absence of radium, from needle to quadrants. This is directly measurable with accuracy. It is nearly proportional to V_n since V is much within 1 per cent. of V_n.

The integral of this equation is, t being the time,

$$V = (\dot{V}_a/K)(1 - KV_c/\dot{V}_a)(1 - e^{-Kt}).$$

If now the needle is left positively charged, but the condenser metals exchanged (commutated), so that the aluminum core is earthed and the shell put in contact with the electrometer (see figure), the equation becomes

$$V' = (\dot{V}_a/K)(1 + KV_c/\dot{V}_a(1 - e^{-Kt}).$$

Let $\kappa = N/K$ and $\kappa' = N/K'$ where K' refers to the normal velocity of positive ions, u. Then if $k = V_c/\kappa\dot{V}_a$, and $k' = V_c/\kappa'\dot{V}_a$, similarly

$$\dot{V} = \dot{V}_a(1 - kN)e^{-Kt}.$$
$$\dot{V}' = \dot{V}_a(1 + kN)e^{-Kt}.$$

If the potential $V = V_\infty$ at $t = \infty$,

$$V_\infty = \kappa\dot{V}_a/N - V_c,\ V_\infty' = \kappa\dot{V}_a/N + V_c,$$

two equations from which both N and V_c may be found, if the limiting potentials V_∞, V_∞', and the electrometer current $\dot{V}_a$ are severally observed. If V_∞ is not obtainable, it may be computed from observations at t and $t_1 = 2t$, as

$$V_\infty = (2V - V_1)/V^2 \text{ and } V_\infty' = (2V' - V_1')/V'^2.$$

Here however there is a difficulty as the curves begin with a double inflection not yet expained. The times $t_1 = 2t$ must therefore be estimated from the observations beyond the double inflections; or the rearward prolongation of the curve for those observations, to meet the time axis. The initial tangents may be found in the same way, but this is not necessary since their values are, respectively,

$$\dot{V}_a(1 - kN) \text{ and } \dot{V}_a(1 + kN).$$

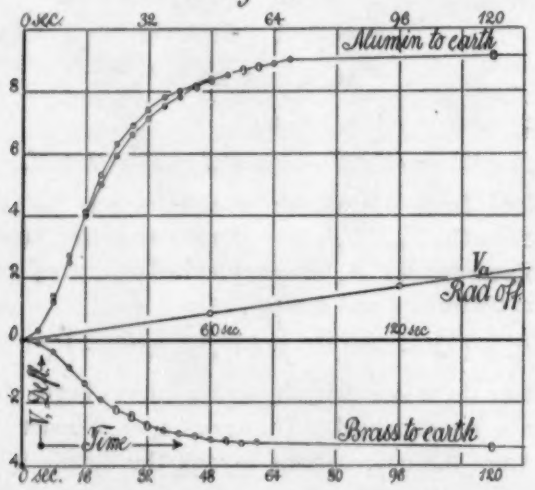
3. *Data: Origin of the Electrometer Current.*—The seat of the chief electromotive force in the electrometer follows from the following data, in which the capacities C, C', C'', Fig. 1, are successively removed. The currents increase in the same ratio as the reduction of capacities, E being that of the electrometer. The data are (potentials in scale parts where 1 cm. is equivalent to .0595 volt), $\dot{V}_a$ being the fall per second:

Capacities.	$\dot{V}_a$ in cm.	$\dot{V}_a$ in Volts.
$C + C' + C'' + E$	.14	.0083
$C' + C'' + E$	.15	.0089
$C' + E$	.58	.0345
E	4.3	.256

The change of voltage throughout the main contours of the curves is almost a linear variation with the lapse of time, except that at the beginning the motion is accelerated from rest as usual; for instance:

Time	0	4	8	12	16	20	24	28	32	60 sec.
V_a	0	.3	1.5	3.7	6.1	8.7	11.0	13.4	15.5	31.0 cm.

Fig 2



4. *Aluminum Tube Charged with Radium Tubelets I.–V.: Data.* —The air in the condenser C is now highly ionized and its voltage becomes appreciable. The data obtained are given in Fig. 2. The needle is positively charged, thus impelling positive charge toward the quadrants. In the four series of data observed the aluminum core of the condenser is twice joined to the electrometer, the brass shell being put to earth (series 1 and 4) and twice commutated (aluminum to earth series 2 and 3). The results are identical except that in series 3 the insulation was perhaps better, or V_a may have changed. The accelerated march of the needle from rest is obvious in both curves and is thus independent of the sign of the limiting voltage, V . It may be mere inertia, but it is of less consequence here because the initial data are not needed in the following computation.

5. *Results: Ionization, N. Voltaic Contact Potential Difference*[1] V_c.—The equations

$$V_c = \kappa \dot{V}_a/N - V \ ,$$
$$-V_c = \kappa' \dot{V}_a/N - V_\infty' ,$$

may now be used to compute N and V_c. The constants are numerically (all in scale parts, 1 cm. equivalent to .0595 volt),

$$\kappa = 36.1 \times 10^6, \quad V_\infty = -3.45, \quad V_a = .142$$
$$\kappa' = 39.7 \times 10^6, \quad V_\infty' = 9.3,$$

Hence

$N = 876{,}000$ ions, either positive or negative,
$V_c = 6.37$ cms., or .376 volts.

[1] [The drift, V_a, which in the above experiments was eliminated by commutation, was eventually traced to a defect in the electrometer. It vanishes on replacing the given instrument by another. Data since obtained for Aluminium-Copper and Aluminium-Zinc condensers showed

Al-Cu, $V_c = .58$ volts,

Al-Zn, $V_c = .06$ volts,

or

Zn-Cu, .52 volts,

a result, however, which varied much with the surfaces, etc.] June, 1909.

BROWN UNIVERSITY,
PROVIDENCE, R. I.

PROC. AMER. PHIL. SOC., XLVIII. 192 N, PRINTED SEPTEMBER 3, 1909.

THE ABSORPTION SPECTRA OF VARIOUS POTASSIUM, URANYL, URANOUS AND NEODYMIUM SALTS IN SOLUTION AND THE EFFECT OF TEMPERATURE ON THE ABSORPTION SPECTRA OF CERTAIN COLORED SALTS IN SOLUTION.

(Plates VII. to XIV.)

By HARRY C. JONES and W. W. STRONG.

(*Read April 24, 1909.*)

(A report on part of the work on absorption spectra that is being carried out with the aid of a grant from the Carnegie Institution of Washington.)

Outline.

I. Experimental Methods.
II. Absorption Spectra of Potassium Salts in Aqueous Solutions.
III. Absorption Spectra of Uranyl Nitrate (under Different Conditions).
  (*a*) In Aqueous Solution.
  (*b*) In the Crystalline State.
  (*c*) As Effected by Dilution.
  (*d*) In Methyl Alcohol.
  (*e*) In Mixtures of Methyl Alcohol and Water.
  (*f*) In Ethyl Alcohol.
  (*g*) As an Anhydrous Salt.
IV. The Absorption Spectrum of Uranyl Bromide, Uranyl Acetate and Uranyl Sulphate.
  (*a*) Uranyl Bromide in Water.
  (*b*) Uranyl Acetate in Water.
  (*c*) The Uranyl Bands of the Acetate.
  (*d*) Uranyl Sulphate in Water.
V. The Absorption Spectrum of Neodymium Chloride in Glycerol.
VI. The Absorption Spectrum of Uranyl Chloride.
  (*a*) In Water.
  (*b*) As an Anhydrous Salt.
  (*c*) The Characteristic Bands in Water.
  (*d*) As Affected by Calcium and Aluminium Chlorides.
  (*e*) In Methyl Alcohol.

(*f*) In Methyl Alcohol with Calcium Chloride.
(*g*) In Methyl Alcohol and Water.
(*h*) In Ethyl Alcohol.
(*i*) The Blue-violet Band.
VII. The Absorption Spectrum of Uranous Salts.
VIII. An Example of the Complexity of the Problem of Explaining the Origin of Spectral Lines and Bands and the Proposed Method of Attacking this Problem.
IX. Effect of Rise of Temperature on the Absorption Spectra of Certain Salts in Aqueous Solutions.
(*a*) Uranous Chloride.
(*b*) Copper Bromide.
(*c*) Chromium, Calcium and Aluminium Chlorides.
(*d*) Uranyl Chloride.
(*e*) Neodymium Salts.
(*f*) Erbium Chloride.
X. Summary.

I. Experimental Methods.

On account of the large number of bands in the absorption spectra of uranium and the rare earth salts, a study of the absorption spectra of these salts is more interesting and more fruitful of results than the study of the absorption spectra of the ordinary colored salt like those of nickel or copper. The absorption spectra have been mapped for potassium ferricyanide, potassium ferrocyanide, potassium chromate, potassium dichromate, the acetate, bromide, chloride, nitrate and sulphate of uranyl in water, of uranyl acetate, nitrate and chloride in methyl alcohol, and of uranyl nitrate and chloride in ethyl alcohol. Beer's law has been tested for these salts as well as the effect of foreign substances on the absorption spectra. The absorption spectra of two uranous salts, the chloride and sulphate, have been photographed and the absorption spectra of neodymium chloride in pure glycerol and in mixtures of glycerol and water have been studied. In this work the methods used by Jones and Uhler[1] and Jones and Anderson[2] have in the main been employed.

The investigations on the effect of changes in temperature on the absorption spectra of solutions have been confined to different concentrations of aqueous solutions of the chloride, nitrate, acetate,

[1] Publication No. 60, Carnegie Institution of Washington.
[2] Publication No. 110, Carnegie Institution of Washington.

sulphate and sulphocyanate of cobalt, the chloride, acetate and sulplate of nickel, the chloride, sulphate and acetate of chromium, chrome alum, the nitrate and bromide of copper, uranous chloride, erbium chloride, the chloride and nitrate of præsodymium, the sulphate, acetate, chloride and nitrate of uranyl and the chloride, bromide and nitrate of neodymium. Spectrograms are made of the absorption spectra for a given concentration of a salt, keeping the thickness of layer constant for every 15° between 0° and 90° C.

To make a spectrogram light from a Nernst glower and from a spark is allowed to pass through the solution that is being investigated. It is then focused upon the slit of a spectroscope—and falling then on a concave grating, the light is spread out into a spectrum on the film upon which it is photographed. The films used were made by Wratten and Wainwright of Croyden, England, and were very uniformly sensitive to light from λ 2100 to λ 7200.

The sectional diagram (Fig. 1) will make the experimental arrangement of the apparatus clearer. N is a Nernst glower which is arranged to slide along the rod AB. P and P' are quartz prisms which are held by a lid L. The prism P is stationary, whereas the prism P' can be moved by the travelling carriage E back and forth through the trough T which contains the solution whose absorption spectrum is being investigated. AB is so inclined that the optical length of the light beam from N to P', P and the concave mirror M shall be constant, whatever the length of the solution between P and P' may be. The greatest length of path PP' used was 200 mm. The hypothenuse faces of P and P' are backed by air films which are enclosed by glass plates cemented to the quartz prisms.

Considerable difficulty was experienced in finding a cement that would adhere to the polished quartz prisms at the higher temperatures. For aqueous solutions baked caoutchouc was found to work fairly well. D is a brass box holding the trough T. D is filled with oil and is placed in a water-bath whose temperature can be varied between 0° and 90° C. The path of a beam of light is then from the Nernst glower (N) or spark to the quartz prism P'. The light is totally reflected from the hypothenuse face of this prism through the solution to P. This prism also has its hypothenuse face backed by an air-film, so that the light is totally reflected upwards to the

concave speculum mirror at *M*. *M* focuses the light on the slit of the Rowland concave grating spectroscope, *G* being the grating and *C* the focal curve of the spectrum. The prism arrangement was designed by Dr. John A. Anderson.

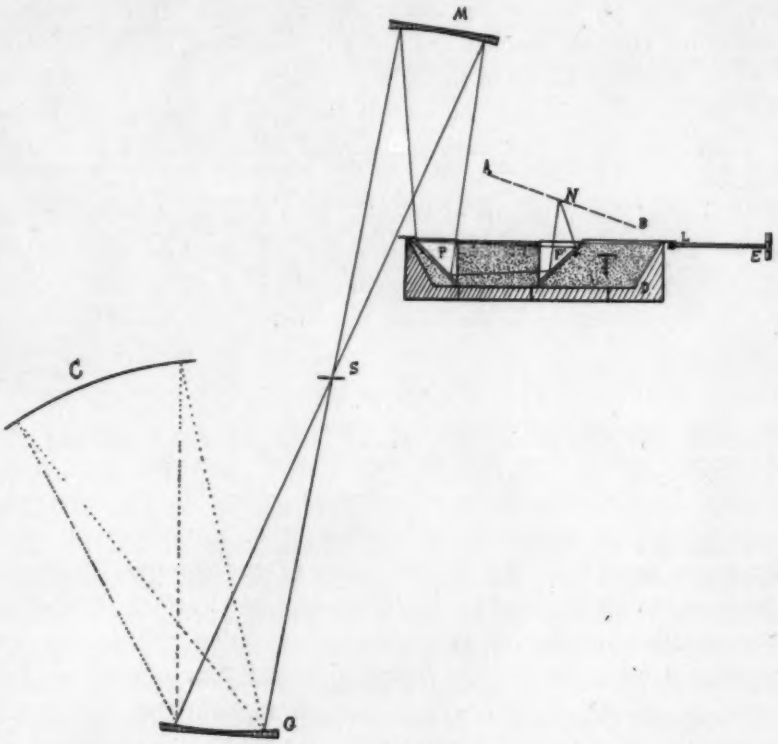


FIG. 1.

This apparatus was found to work very well for aqueous solutions. Some evaporation took place at the higher temperatures, but distilled water was added in proper quantity and mixed with the solution so as to keep the concentration constant. By using troughs of different lengths it was possible to vary the length of salt solution through which the light beam passed from 1 to 200 mm. One inconvenience was experienced at low temperatures; moisture would sometimes condense upon the exposed prism faces. To overcome this an air blast was directed upon these faces and this helped very materially to prevent the condensation of moisture.

II. Absorption Spectra of Potassium Salts in Aqueous Solutions.

Most potassium salts in solution are colorless, and for this reason it is considered that the potassium atoms do not themselves absorb any light in the visible portion of the spectrum. Several colored potassium salts are known and the color of these is due in some way to the other atoms in the salt molecules. In the present work the absorption spectra of potassium ferricyanide, potassium ferrocyanide, potassium chromate and potassium dichromate have been studied.

Using a 3 mm. length of solution of potassium ferricyanide in water we find that for a normal concentration there is complete absorption of all the shorter wave-lengths of light beyond λ 4800. As the concentration is decreased the edge of transmission moves continually towards the violet. It should be noticed that the region between complete absorption and complete transmission for the more concentrated solutions is quite narrow, being less than 40 Ångström units; thus making solutions of this salt quite good screens for absorbing light. Continually decreasing the concentration we reach a 0.0156 normal solution, when a transmission band begins to appear. For a certain range of concentration there appears an absorption band in the region λ 4200. Further decrease in concentration results in increasing transmission throughout the violet and ultra-violet. For dilutions greater than 0.00195 normal there is almost complete transmission throughout the ultra-violet. Very faint bands appear in the regions λλ 2500 to 2600, λλ 2950 to 3050 and λλ 3200 and 3250.

Several spectrograms were made, keeping the product of concentration and depth of solution layer constant. In this case the spectrograms will be identical if Beer's law holds. Beer's law was found to hold according to this method of testing within the ranges of concentration over which the spectrum was mapped.

The absorption of aqueous solutions of potassium ferrocyanide was investigated in the same way. A half-normal solution 3 mm. deep shows that all light of shorter wave-length than λ 3950 is absorbed. Keeping the depth of layer the same, it is found that

with decrease in concentration the transmission gradually moves towards the ultra-violet, and for dilutions greater than 0.0078 normal there is transmission throughout the whole spectrum. Beer's law was found to hold.

A 2-normal aqueous solution of potassium chromate 3 mm. in thickness, shows complete transmission of wave-lengths greater than λ 4950. Decreasing the concentration causes the transmission to move gradually towards the violet and for a 0.01 normal solution a transmission band appears at λ 3100, or, in other words, there appears an absorption band whose center is about λ 3700. As the concentration decreases this absorption band fills up, the violet edge of the transmission band gradually pushes out into the ultra-violet, and for dilutions greater than 0.0005 normal there is complete transmission throughout the spectrum. Beer's law was found to hold for potassium chromate throughout the above ranges of concentration, except in the more concentrated solutions between 2 normal and 0.25 normal.

Potassium dichromate in water was found to have a much greater absorbing power than the solutions previously described. A one-third normal concentration absorbed all wave-lengths shorter than λ 5350. As the concentration is decreased the transmission extends farther and farther out into the violet. For a 0.0026 normal concentration a transmission band appears in the violet, thus giving an absorption band whose center is about λ 3800. As the concentration is further decreased transmission becomes greater and greater in the violet and ultra-violet, and is practically complete for a 0.0006 normal concentration. Beer's law has been tested between the above ranges of cencentration and has been found to hold.

In photometric measurements of Beer's law, the equation defining the quantities to be measured is:

$$J = J_0 \, 10.^{-Acl}$$

J_0 is the intensity of the light that enters the solution (neglecting any loss due to reflection), J the intensity of the light as it leaves the solution, c the concentration in gram molecules of the salt per liter of solution, l the thickness of layer and A a constant if Beer's law holds. Strictly speaking the above equation holds for mono-

chromatic light. For ordinary white light one would have to integrate this equation over the range of wave-lengths used. The equation would then have the form

$$J = J_0 \int_{\lambda_1}^{\lambda_2} e^{-\beta t c} d\lambda.$$

The quantity β is called the index of absorption and A the molecular extinction coefficient. If the absorption is proportionately greater in the more concentrated solutions, then Beer's law fails and A decreases inversely as the concentration.

From photometric measurements Settegast[3] and Sabatiér[4] conclude that the absorption spectrum of potassium dichromate is the same as that of chromic acid, and that the absorption spectrum of potassium chromate is entirely different. This is corroborated by the present work. Settegast finds that Beer's law does not hold for potassium chromate and potassium dichromate, the coefficient A decreasing with increasing concentration. Grünbaum[5] finds the following values of A and ϵ where $\epsilon = c/A$.

	Potassium Dichromate.	
	Value of A.	Value of A.
λ	$c = .034$	$c = .0034$
509	62.4	58.0
521	28.7	26.2
538	7.24	6.2

It will be seen that the deviation here from Beer's law is in the opposite direction from that of Settegast. Grünbaum finds that ϵ and therefore A depends on the depth of layer.

An example will be given where the same concentration was used and different depths of the solution.

λ	Values of ϵ for $c = .0034$		
	25 cm. layer.	10 cm. layer.	5 cm. layer.
521	.0758	.0818	.0884
521	.0761	.0830	.0897

Our work indicates that Beer's law holds for all small concentrations and usually the deviations for concentrated solutions is very

[3] *Wied. Ann.*, 7, pp. 242–271, 1879.
[4] *C. R.*, 103, pp. 49–52, 1886.
[5] *Ann. d. Phys.*, 12, pp. 1004, 1011, 1903.

small. Of the potassium salts above described, only potassium chromate between 2 normal and 0.25 normal showed any considerable deviation from Beer's law, and in this case the absorption of the concentrated solution was greater than would be expected if Beer's law held by about 40 Ångström units.

The present method is a very good qualitative test of Beer's law, and gives the results for each wave-length, whereas most photometric methods only give integrated results over a more or less wide region of wave-lengths.

III. Absorption Spectrum of Uranyl Nitrate under Different Conditions.

There are two groups of uranium salts, the uranyl salts containing the UO_2 group, and the uranous salts. The uranyl salts in solution are yellow and usually crystallize from aqueous solutions with a certain amount of water of crystallization; for example, at ordinary temperatures uranyl sulphate crystals have the composition $UO_2(SO_4)_23H_2O$. The uranous salts are intensely green and are very unstable, oxidizing very easily to the uranyl condition. Uranous sulphate crystals have the composition $U(SO_4)_29H_2O$.

(*a*) *Uranyl Nitrate in Aqueous Solution.*

The spectrum of uranyl nitrate in water is a typical example of the uranyl salts. Using a depth of solution of 3 mm. its absorption spectra was investigated between concentrations of 1.5 normal to 0.0234 normal. For the 1.5 normal solution the absorption consists of a band in the blue-violet and absorption throughout the ultra-violet portion of the spectrum. As the concentration decreases the blue-violet band fills up with transmission, and the ultra-violet absorption is pushed farther and farther out into the ultra-violet. The blue-violet band is practically gone at a concentration of 0.5 normal, and there is almost complete transmission throughout the ultra-violet for concentrations less than 0.02 normal.

During these changes in concentration a large number of bands about 50 Ångström units wide make their appearance. Near the edge of an absorption band these bands are relatively quite clear.

As the absorption edge recedes from the uranyl bands, the general transmission is so great as almost to entirely obscure them.

A, Plate I, represents the absorption spectra of an aqueous solution of uranyl nitrate of different depths of layer. The narrow and rather weak bands shown here are the uranyl bands. Twelve of these bands have been photographed. Starting at the band of longest wave-length they shall be designated by the letters *a, b, c, d,* etc. On account of the irregularity of the distribution of light in the spark spectrum and the small intensity of the uranyl bands, the Nernst glower was used as the source of light in the ultra-violet, and long exposures were made. A screen was used that cut out all wave-lengths greater than λ 4200. *A* represents a typical spectrogram of this kind. Starting with the spectrum strip at the top, the concentrations were 1.5 N, 1.1255 N, 0.75 N, 0.5 N, 0.375 N, 0.25 N, and 0.1875 N. The slit width was 0.08 mm. and the current through the Nernst glower 0.8 amperes. The spectra of wave-lengths greater than λ 4300 represent the absorption of a depth of layer of 15 mm.; the spectra of shorter wave-lengths represent the absorption of a depth of layer of 3 mm. The upper spectrum strip represents then the absorption spectrum of a 1.5 normal solution of uranyl chloride 15 mm. thick, exposure being made 1 min. to the Nernst glower. It will be seen that the uranyl *a* band comes out very strongly. The screen was then placed in the path of light and exposure of 5 minutes made to the violet and ultra-violet beyond λ 4300; a solution of uranyl nitrate of 1.5 normal concentration and 3 mm. depth of cell being in the path of the beam of light. This amount of uranyl nitrate absorbed practically all the light in this region. A very short exposure was afterwards made to the spark in the region λ 2600, in order to get a comparison spark spectrum in this region, so that the wave-lengths of the uranyl bands could be measured.

Throughout this work a comparison spark spectrum usually containing the very strong line λ 2478.8 was photographed on each spectrum strip. In measuring the uranyl bands all measurements were made from this line as a standard, and although the absolute wave-lengths of the uranyl bands may not be correct to within 20

Ångström units, yet their relative accuracy is probably correct to within less than 10 Ångström units for the finer bands.

The second spectrum strip from the top represents in the long wave-length end of the spectrum the absorption of a 15 mm. solution of a 1.125 normal solution of uranyl nitrate exposed 1 min. to the Nernst glower. The *a* band appears, although not nearly as intense as in the spectrum strip above. The region of shorter wave-lengths beyond λ 4300 represents the absorption of a 3 mm. depth of layer of a 1.125 normal concentration exposed 5 min. to the Nernst glower. A very faint transmission is shown in the region λ 3700. The ultra-violet line λ 2478.8 is shown in the comparison spark spectra. The other spectrum strips were made in a similar manner, using the concentrations given above.

By this method of exposing two new bands were detected in the ultra-violet. In aqueous solutions the intensities of the bands are much the same. In other solvents however and for other uranyl salts, the relative intensities of the bands change very greatly. In uranyl nitrate crystals the bands are even more closely related to each other than in aqueous solutions. The longer the wave-length of the band the more intense and wider it is as a rule. The position of the long wave-length bands in the orthorhombic uranyl nitrate crystals $UO_2(NO_3)_2 6H_2O$ is the same as the position of the bands for an aqueous solution. The wave-lengths of the bands are as follows:

	a	*b*	*c*	*d*	*e*	*f*	*g*	
Water Sol.	4860	4720	4540	4380	4290	4150	4020	Deussen.
Water Sol.	4870	4705	4550	4390		4155	4030	Jones and Strong.
Crystals	4870	4705	4500–4565	4405	4275	4170	4050	

	h	*i*	*j*	*k*	*l*	
Water Sol.	3870	3790	3690			Deussen.
Water Sol.	3905	3815	3710	3605	3515	Jones and Strong.
Crystals	3935	3830	(3720?)	3600		

In the original film from which *A*, Plate I, was made all these bands except *d* could be very distinctly seen. The bands of longer

wave-length are slightly wider. The *i* band is considerably weaker than its neighboring bands.

(*b*) *Absorption Spectrum of Uranyl Nitrate Crystals.*

In the aqueous solution there is no sign that the bands can be broken up. In the crystal spectrum this is not the case. The *a* band is narrow. The *b* band is also very narrow, about 15 Ångström units wide. A very faint band appears about λ4650. The *c* band, on the other hand, is very wide, about 70 Ångström units, and is probably double. The *d* band is about 50 å. u. wide, and the *e* band is about 70 Ångström units wide and appears double. The *f* band is the most intense and is about 40 Å. u. wide. The bands *g, h, i* and *j* keep decreasing in intensity respectively. The above description is of a spectrogram taken of a crystal in Canada balsam, and of course the width of the bands varies with the time of exposure and various other things. The above spectrogram showed many details, however, that other spectrograms did not. It will thus be seen that the *a, b, c, d, j* and *k* bands of the solution agree fairly well with those of the crystal, and that the crystal bands *f, g, h* and *i* are shifted towards the red with reference to the bands in the aqueous solution.

(*c*) *Effect of Dilution upon the Uranyl Bands.*

The effect of dilution on the position and intensity of the blue-violet, the ultra-violet and the uranyl bands of the acetate, nitrate and sulphate of uranyl in water was tried. The absorption spectra of solutions of about 1 normal and 3 mm. depth of cell was photographed along by the side of the absorption spectra of the same salts of 0.008 normal concentration and 380 mm. depth of layer. The absorption consisted of the blue-violet band, the ultra-violet band and the *a, b, c, i, j* and *k* bands. Between the blue-violet and ultra-violet bands there was the transmission band containing *i, j* and *k*. For each of the three salts this transmission band was much weaker for the dilute solution, whereas in the cases of the sulphate and nitrate the long wave-length transmission edge of the blue-violet band was stronger for the more dilute solution. The opposite was true of the acetate solution. In the dilute solution of

the acetate the bands were more intense than for the more concentrated solution. There was no noticeable change in the position of the bands. Neither the intensity nor the position of the uranyl nitrate or the uranyl sulphate bands was changed by the above dilution.

A more detailed study was made as to whether Beer's law holds for uranyl nitrate and for the other uranyl salts. The method of taking the spectrograms is the same as that used for the potassium salts.

Beer's law was found to hold for dilute solutions of uranyl nitrate in water. When the concentration is greater than .5 normal the absorption is greater than it should be if Beer's law held.

(*d*) *Uranyl Nitrate in Methyl Alcohol.*

In methyl alcohol the general appearance of the absorption is very similar to that of the aqueous solution; the blue-violet, the ultra-violet, and uranyl bands appearing under the same general conditions that they appear for aqueous solutions. There is a very marked deviation from Beer's law for the more concentrated solutions, however; the absorption of concentrated solutions being greater than it would be if Beer's law held. The positions of the bands are quite different from the positions of the uranyl bands of the aqueous solution, or of the crystals, as shown by the following values:

	a	*b*	*c*	*d*	*e*	*f*	*g*	*h*	*i*
λ	4930	4760	4610	4455	4325	4190	4070	3965	3855

(*e*) *Uranyl Nitrate in Mixtures of Methyl Alcohol and Water.*

In the previous work of Jones and Anderson[6] it was found that in some cases (for example neodymium chloride) a salt in water had a different set of absorption bands compared with the same salt in another solvent as, *e. g.*, methyl alcohol.

When the salt is dissolved in mixtures of these two solvents, say methyl alcohol and water, it was found that as the amount of one solvent, methyl alcohol for instance, decreased the methyl

[6] Publication No. 110, Carnegie Institution of Washington.

alcohol bands of the salt decreased in intensity, but did not change their position in the spectrum. At the same time the water bands of the salt became more intense. In the present work it is shown that the uranyl nitrate bands in pure water and in pure methyl alcohol occupy different positions. The problem to be investigated is to find out whether in mixtures of water and methyl alcohol, the uranyl bands will show a gradual shift, or whether the methyl alcohol uranyl bands and the water bands will both exist together; their relative intensities being proportional to the relative amounts of methyl alcohol and water. It was found that the two sets of bands exist together and that the methyl alcohol bands decrease in intensity quite rapidly with increase of water. The blue-violet band showed marked changes until the amount of water reached about 20 per cent. In this work the amount of uranyl nitrate in the path of the light was kept constant, and the only variable was the relative amounts of methyl alcohol and water. The above would indicate that uranyl nitrate in water is "hydrated" and in methyl alcohol it is "alcoholated." The above data indicate that the effect of "hydration" is much more persistent than that of "alcoholation." It is quite possible that this is due to a greater number of water molecules producing the hydration than there is methyl alcohol molecules taking part in alcoholation.

(*f*) *Uranyl Nitrate in Ethyl Alcohol.*

The absorption of uranyl nitrate in ethyl alcohol was mapped and the general characteristics were found to be the same as for the water and methyl alcohol solutions. A new band was found at λ 5200 which was about 50 Ångström units wide. All the uranyl bands were very faint and wide and therefore difficult to measure. Beer's law showed deviations similar to those found for the methyl alcohol solution. On account of the diffuseness of the bands no spectrograms were made of mixtures of water and ethyl alcohol. Following are approximately the positions of a few of the bands:

	a	*b*	*c*	*d*	*e*	*f*	*g*	*h*	*i*
λ	5000	4800	4630	4475	4325	4180	4080	3970	3875

(*g*) *Absorption Spectrum of Anhydrous Uranyl Nitrate.*

When it was first discovered that the uranyl nitrate "water" bands were all shifted to the violet with reference to the bands of the other uranyl salts in water, as well as with reference to the uranyl nitrate bands in other solvents, it was thought that possibly it was more hydrated than the other salts in solution. The uranyl salts crystallize from water solutions at ordinary temperatures with the following composition: $UO_2(NO_3)_2.6H_2O$, $UO_2SO_4.3H_2O$, $UO_2(CH_3COO)_2.2H_2O$, and $UO_2Cl_2.H_2O$. This fact would favor the supposition that in solution the nitrate might be more hydrated than the other salts. The fact that the absorption of the aqueous solution of the nitrate and the crystallized salt was very much the same as far as the positions of the uranyl bands is concerned, also seemed to favor this view.

In this connection it was considered important to examine the absorption spectrum of the anhydrous uranyl nitrate. The salt was powdered and placed in a closed glass tube just above the slit of the spectroscope. The light of a Nernst glower was then focused upon the surface of the salt nearest the slit and an exposure of about three hours made. In this way we examine light that has penetrated a short distance into the powder and is then diffusely reflected.

The absorption spectrum was found to consist of quite a large number of bands, which seem quite different in many respects from those of the solution. The following are the approximate wave-lengths: λλ 4800, 4650, 4500, 4420, 4360, 4280, 4180 (broad), 4060 (broad), 3950 (broad), 3820 (broad), 3700 (narrow) and 3600 (narrow). The bands marked broad are from 50 to 60 Ångström units wide and the narrow bands about 20 Ångström units. If the first band is the *a* band, then the bands of the anhydrous salts are to the violet of the corresponding bands of the crystals and of the solution. If it is the *b* band the opposite is the case. On account of the smallness of the intensity of the bands it could not be settled whether λ 4800 is the *a* or the *b* band. Further investigation of this point will be made.

There are two difficulties to the above theory, difficulties for

which no explanation so far has been suggested. In the work on the effect of rise of temperature on the absorption spectrum it was found that the uranyl nitrate bands did not shift to the red. On the other hand, the uranyl sulphate and uranyl chloride bands were shifted to the red under the same conditions. (In these cases aqueous solutions were investigated.) If the uranyl nitrate bands owe their position to a large amount of hydration it would be expected that with rise in temperature they would be shifted towards the red more than the bands of the sulphate and chloride. Another difficulty is that of the effect of dilution. The greater the dilution the greater the dissociation, and, therefore, according to the theory of Arrhenius for very dilute solutions the UO_2 group should exist in the ionic condition and the absorption spectrum of all the salts should be the same, *i. e.*, the uranyl bands should then occupy the same positions independent of the kind of salt. No effect of this kind is to be noticed, as was shown above under the division describing the effect of dilution. It is intended to use much more dilute solutions in the future.

IV. The Absorption of Uranyl Bromide, Uranyl Acetate and Uranyl Sulphate.

(*a*) *Absorption Spectrum of Uranyl Bromide in Water.*

The absorption spectrum of uranyl bromide in water was mapped and found to be very similar to that of the nitrate. The ultra-violet, blue-violet and uranyl bands appear and are affected in the same manner as the same bands of the nitrate. Beer's law was found to hold. The uranyl bands were found to be much wider and more diffuse than in the case of the aqueous solution of the nitrate. The following are their approximate positions:

a	*b*	*c*	*d*	*e*	*f*
4880	4720	4560	4450	4280	4160

(*b*) *Uranyl Acetate in Water, Beer's Law.*

A spectogram was made to test whether Beer's law holds for an aqueous solution of uranyl acetate between the concentrations 0.25 normal and 0.031 normal. The spectrogram showed that there was a very great deviation from the law, and in the opposite direc-

tion to any deviation hitherto found either in this work or in that of Jones and Anderson or Jones and Uhler. The absorption of the more dilute solutions was found to be proportionately much greater than for the more concentrated solutions. A similar run was made for a solution of the acetate in methyl alcohol and a deviation from Beer's law in the same direction was found, although the amount was not so great in this case.

(*c*) *The Uranyl Bands of the Acetate.*

The following table gives the approximate wave-lengths of the uranyl bands of the acetate in water, in methyl and as the anhydrous powder.

Bands of Uranyl Acetate.

	a	*b*	*c*	*d*	*e*	*f*	*g*	*h*	*i*
In Water	4910	4740	4595	4455	4310	4160	4070	3970	3830
In Methyl Alcohol	4900	4770	4600	4460	4320	4200	4090		
As Anhydrous Salt.	4910	4760	4610	4460	4330	4190	4070	3980	

From this table it seems that the positions of the bands of the acetate under these different conditions is about the same.

(*d*) *Absorption Spectrum of Uranyl Sulphate.*

The mapping of the absorption spectrum of uranyl sulphate in water showed that it was very much like that of the nitrate in water. As in the case of the nitrate the *i* band was much weaker than the adjacent bands. Beer's law was found to hold. The addition of a large amount of sulphuric acid was found to make the uranyl bands much sharper, but not to cause them to shift. Much more work will be done on the effect of strong acids on the uranyl bands. The following gives the wave-lengths of the sulphate bands:

a	*b*	*c*	*d*	*e*	*f*	*g*	*h*	*i*	*j*	*k*	*l*
4900	4740	4580	4460	4330	4200	4070	3970	3850	3740	3630	3530

V. The Absorption of Neodymium Chloride in Glycerol and Mixtures of Glycerol and Water.

The absorption spectrum of a glycerol solution of neodymium chloride is much like that of the aqueous solution in its general

PROC. AMER. PHIL. SOC. XLVIII. 192 O, PRINTED SEPTEMBER 3, 1909.

characteristics, but when proper concentrations are used so as to bring out the fine bands it is found that the two spectra are entirely different. For example, the aqueous solution shows a very fine band at λ 4274. In the glycerol there is a band that on first sight appears exactly identical with this 4274 band. However, its wave-length is about λ 4287, and it has two extremely fine components on each side, one at λ 4273 and one at about λ 4300. The same is true throughout the spectrum.

In general, in mixtures of water and glycerol the appearances indicate that there are "glycerol" bands and "water" bands and as the amount of one solvent is increased, so are the bands corresponding to this solvent increased in intensity. Herein lies a very large field for investigation and considerably more work is being carried on here along these lines. The above described spectrum of the glycerol solution of neodymium indicates that glycerol has a a very great influence upon the vibrations of the electrons within the neodymium atom—and that this is due to a kind of "atmosphere" of glycerol about the neodymium atom. Jones and Anderson showed that alcohol has a similar effect, and that the "alcohol" bands were much less persistent than the water bands. Further work is being done upon the relative persistence of "water," "alcohol" and "glycerol" bands; also on the effects of foreign substances and rise of temperature upon these bands, both in the pure solvent and for mixtures of solvents.

VI. Absorption Spectrum of Uranyl Chloride.

The absorption spectrum of uranyl chloride was mapped for an aqueous solution, a methyl alcohol solution, an ethyl alcohol solution, a mixture of methyl alcohol and water, a mixture of methyl alcohol and calcium chloride, and a mixture of water and aluminium chloride.

(*a*) *The Absorption Spectrum of Uranyl Chloride in Water.*

The absorption spectrum of uranyl chloride in water was found to be very similar in general to that of the other uranyl salts. The uranyl bands were less sharp than the bands of the nitrate and sulphate in water. The wave-lengths of a few of the bands are as follows:

a	*b*	*c*	*d*	*e*	*f*	*g*
4920	4740	4560	4460	4315	4170	4025

(*b*) *Absorption Spectrum of Anhydrous Uranyl Chloride.*

The absorption spectrum of the anhydrous uranyl chloride was photographed in the same way as that of the anhydrous nitrate. The bands differ considerably from the bands of the aqueous solution, and one cannot tell very well whether they are identical with the corresponding *a*, *b*, *c*, etc., bands of the solution or not. Their wave-lengths are approximately as follows: λλ 4950 (narrow), 4860, 4765, 4700, 4615, 4540, 4460, 4320, 4290, 4160, 4050 and 3940.

(*c*) *The Characteristic Bands of Uranyl Chloride.*

In addition to the bands already described, uranyl chloride has several remarkably fine bands in the green. These bands are not more than 5 Ångström units wide and were first seen on spectrograms taken upon Whatten and Wainwright red sensitive films. They appear only for aqueous solutions, and the addition of calcium chloride or aluminium chloride causes them to disappear. They do not appear in alcoholic solutions. Aqueous solutions of uranyl sulphate show them very faintly. The wave-lengths are approximately as follows:

λλ 5185, 5200, 6000, 6020, 6040 and 6070.

These bands have never hitherto been noticed as absorption bands. H. Becquerel[1] gives quite a full set of measurements of the phosphorescent bands of various uranyl salts at room temperature and at the temperature of liquid air. Among the bands given for the double chloride of uranyl and potassium at room temperature are λλ 6070 to 6040, and λλ 5220 to 5193. Whether these correspond to the above absorption bands is quite difficult to say. Further work is being done in this direction.

(*d*) *Uranyl, Calcium and Aluminium Chlorides in Water.*

Spectograms were taken of aqueous solutions of a constant concentration of uranyl chloride to which varying amounts of calcium

[1] *C. R.*, t. 101, p. 1252, 1885; pp. 459 and 621, 1907.

chloride were added. The addition of calcium chloride causes the ultra-violet, the blue-violet band and the uranyl bands to widen generally. The effect upon the uranyl bands is however, very small. The effect of aluminium chloride, however, is very great. The two narrow and faint bands at λ 5200 only appear in the pure aqueous solution of uranyl chloride. The *a* band in the aqueous solution is about 60 Ångström units wide, and is almost as intense as the *b* band. The addition of aluminium chloride causes the band to become quite narrow, about 25 Ångström units wide. A slight addition of alminium chloride decreases the intensity of the band very considerably. Further increases in the amount of aluminium has very little effect. The addition of alminium also causes the bands to shift to the red; the shifts in some instances amounting to 25 Ångström units. The *b* and *c* bands have their intensity very greatly increased by the addition of aluminium chloride; and by making the solution about 2 normal of aluminium chloride these bands are shifted about 30 Ångström units to the red compared with the same bands for the pure uranyl chloride solution. The *d*, *e*, *f*, *g* and *h* bands are also increased in intensity, but are but very slightly shifted to the red. The *d* and *e* bands are widened so that they practically form a single band.

(*e*) *Absorption Spectrum of Uranyl Chloride in Methyl Alcohol.*

In the absorption spectrum of uranyl chloride in methyl alcohol the *a*, *b*, *c*, *d*, *e*, *f*, *g*, *h*, *i*, and *j* bands all appear, the *b* and *c* bands being the largest and most intense. The following are the approximate wave-lengths of the bands:

	a	*b*	*c*	*d*	*e*	*f*	*g*	*h*	*i*	*j*
Uranyl Chloride in Methyl Alcohol	4930	4760	4590	4465	4345	4220	4090	3980	3860	3760
Uranyl Nitrate in Methyl Alcohol	4930	4760	4610	4460	4325	4190	4070	3970	3855	
Uranyl Acetate in Methyl Alcohol	4900	4770	4600	4460	4320	4200	4090			

It is seen from the above table that the uranyl bands of these three salts in alcohol occupy almost exactly the same positions.

(*f*) *Absorption Spectrum of Uranyl Chloride and Calcium Chloride in Methyl Alcohol.*

In the solution of uranyl chloride in methyl alcohol the *d* and *e* bands are very diffuse, but are entirely separate. By adding calcium chloride a very peculiar phenomenon takes place. The *d* and *e* bands come together and as far as one can tell form a single band. At the same time the *f*, *g* and *h* bands shift to the red. For a solution containing a .9 normal solution of calcium chloride one finds that the *b* and *c* bands have practically remained in the same position, the *d* and *e* bands have merged into one and the *f*, *g*, *h* bands have moved to approximately λλ 4260, 4120 and 4010 respectively. The *de* band is approximately at λ 4420.

(*g*) *Absorption Spectrum of Uranyl Chloride in Methyl Alcohol and Water.*

A spectrogram was made of a solution of uranyl chloride of constant concentration in mixtures of methyl alcohol and water. A small addition of water causes a considerable decrease in the absorption power of the uranyl chloride. When the amount of water has reached about 16 per cent. very little further change is produced by further increasing the amount of water. The most important effect of the addition of water is the effect upon the uranyl bands. For a pure alcoholic solution the *a* and *b* bands appear; the *b* band being quite intense. Adding water causes *a* and *b* to both decrease in intensity and apparently to shift towards the violet. A spectrogram of smaller concentration shows the *a*, *b*, *c*, *f*, *g*, *h* and *i* bands; the solution containing 8 per cent. water the *b*, *c*, *d*, *e*, *f*, *g*, *h*, *i* and *j* bands; the 16 per cent. water solution *b*, *c*, *d*, *e*, *f*, *g*, *h*, *i* and *j*; the 24 per cent. aqueous solution shows all these bands greatly weakened, and in solutions containing a greater amount of water practically only the *b* and *c* bands appear, and these are very diffuse. The general effect upon the positions of the bands is quite remarkable, the *b* and *c* bands apparently being shifted to the violet with increase of water, whereas the ultra-violet bands appear to be shifted towards the red.

(*h*) *Absorption of Uranyl Chloride in Ethyl Alcohol.*

The absorption spectrum of uranyl chloride in the ethyl alcohol shows the uranyl bands quite strongly, although they are less intense than for the methyl alcohol solution. A very interesting resemblance has been found for the various uranyl bands of different mixtures. The absorption spectrum of a solution of uranyl chloride in ethyl alcohol has been found to be almost the same as that of a methyl alcohol solution of uranyl chloride containing a 0.9 normal concentration of calcium chloride or an aqueous solution of uranyl chloride and a 2 normal concentration of aluminium chloride.

The positions of the uranyl bands for the ethyl alcohol solution was approximately:

	a	*b*	*c*	*d*	*e*	*f*	*g*	*h*	*i*
λ	4900	4750	4580	4400	4400	4250	4100	3980	3860

The relation above mentioned comes out much better in comparing the spectrograms. The values of wave-lengths thus far given does not bring this out very well on account of the difficulty of making measurements. Much more work is to be carried on along this line, and the measurements above given are to be considered as more or less of a preliminary character.

(*i*) *The Blue-Violet Band.*

Under the various changes above noted, *i. e.*, of changing the acid radicle, of changing the solvent and of adding foreign substances, the position at which the blue-violet band faded away was approximately λ 4200. This is rather unexpected when we consider the very considerable effects which are produced upon the finer bands.

VII. Absorption Spectrum of Uranous Salts in Solution.

It is quite well known that by reduction the yellow uranyl salts are changed to the intensely green uranous salts. In the present work this reduction was accomplished by adding the same acid to the solution that corresponded to the anion of the salt and then putting in a metal that would produce a colorless salt.

The absorption spectrum of uranous sulphate and uranous

chloride in water was found to be very similar. The absorption of the shorter wave-lengths was complete under the conditions used. The following are the approximate positions of some of the bands: λλ 6700, 6500, 6300, 5480 and 4900. The 6500 band was the strongest one of all, and upon increasing the depth of cell this band widened out so as to unite with the bands λλ 6700 and 6300, forming an absorption band covering hundreds of Ångström units. This is a very characteristic property of many of the uranous bands, that of widening out so as to include a very large portion of the spectrum. The uranyl bands do not change in width very greatly on increasing the depth of cell.

Besides the bands described above uranous chloride shows bands at λλ 4600, 4770 and 4970.

The absorption spectrum of uranous chloride in methyl alcohol was found to differ very much from that of the aqueous solution. The bands at λλ 4600 and 4780 appeared, closely resembling the water bands at the same position. The band λ 4970 in water was broken up into two bands in methyl alcohol at λλ 4930 and 5030. In the alcohol a very broad band appeared at λ 5300, which does not appear at all in the water solution. The band at λ 5580 is very similar to the water band. Weak and broad bands appear at λλ 6150, 6300 and 6480, and a strong band at λ 6750. As the depth of the alcoholic solution is increased the widening of the bands is very different from the widening of the bands of the aqueous solution.

The absorption spectrum of a mixture of calcium chloride and uranyl chloride in water was found to be very similar to that of the pure uranyl water solution. Much further work along the above lines is being carried on.

VIII. An Example of the Complexity of the Problem of Explaining the Origin of Spectral Lines and Bands and the Proposed Method of Attack.

It is a fact that investigations upon the spectral emission and absorption of bodies has been far less fruitful in extending our knowledge of the structure of the atom than had been expected.

This is largely owing to the almost infinite complexity of the structure of the atom and our general ignorance of the forces that exist there. Probably the best known example is that of the uranyl group which we have been describing. Let us consider the spectral vibrations that can be produced by components that exist or may be produced from the apparently simple UO_2 group: (1) We have the absorption spectrum described above. At low temperatures most of these bands break up into much finer bands. (2) The uranyl salts under various methods of excitation emit a phosphorescent spectrum of a large number of rather fine bands throughout the visible region of the spectrum. It is quite possible that this spectrum is intimately connected with that of the absorption spectrum. (3) We have the absorption spectrum of the uranous salts which has been described above. This spectrum has been probably caused by the change of valency of the uranium atom. Uranium is known to form quite a large number of oxides and it is quite possible that for each valency of the uranium we have a characteristic spectrum. (This also is being investigated.) It is also quite probable that at low temperatures those spectra would consist of quite fine bands. (4) We have the spark spectrum and the absorption spectrum of oxygen, and (5) that of ozone, which bears no relation to that of oxygen. (6) There is the exceedingly complex spark spectrum of uranium consisting of thousands of fine lines and also (7) the complex arc spectra. From radioactive experiments it is known that uranium is continually breaking down into ionium. (8) Ionium possesses the properties of a chemical atom and most likely has a spectrum of its own. This would make eight spectra. (9) Ionium breaks down into the radium and radium has a very characteristic spark spectrum, as does also (10) the radium emanation. During the various radioactive transformations several α products are emitted with a velocity almost as great as that of light. It is probable that these particles are moving with very great velocities in the uranium atom under ordinary conditions. (11) The α particles are known to be charged helium atoms and therefore under proper excitation would give the helium spectrum. The radium emanation breaks down into Radium A, B, C, D, E and F. These products behave like chemical elements

and probably have characteristic spectra. (12) The final product is lead, which has a very complex spark and arc spectra. During these transformations several electrons have been thrown off from the various products with enormous velocities. In a very large number of the above spectrum lines the Zeeman effect indicates the presence of negative electrons and charged doublets.

We thus see what an extremely complex system UO_2 must be and it might seem almost hopeless to entangle the mystery of its various spectra. At present we know that the arc and spark spectra problem is very complex and that we have very few methods of producing any changes in it. Practically the only method of changing the frequency of these vibrations is by applying a very powerful magnetic field or great pressure and these changes in the frequency are very small. One very important result, however, has been accomplished by Kayser, Runge, Wood and others. This work consists in separating spectrum lines into various series. A series of lines are those whose intensity and Zeeman effect vary in the same way when the conditions outside the atom are changed. The work of Wood is very important and shows that spectrum lines are due to different systems of vibrators inside the atom. By using monochromatic light of different wave-lengths he has been able to excite diffirent series of lines which constitute altogether the fluorescent spectrum of the element.

Present theories of the atom usually regard the electrons and other vibrators that are the sources of arc and spark lines as being far within the atom and as affected by external physical conditions only under very special circumstances. Stark believes that these electrons may rotate in circular orbits, the locus of the centers of these orbits being a closed curve, say a circle. This system will be equivalent to a positive or negative charge according to the sense of rotation of these electrons. These electrons we will call ring electrons. Supposing these systems to be positive charges, it will require electrons to neutralize these charges. Several of these neutralizing electrons may be in the outer parts of the atom and under certain conditions might be knocked off from the atom. These easily removable electrons will be called "valency" electrons, and can exist under different conditions of "looseness" of con-

nection with the atom. Most of the neutralizing electrons will probably lie far within the atom. For instance, we would expect that in the uranium atom the charged helium atoms are neutralized by negative electrons.

Our theory is that the finer absorption bands of such salts as neodymium, erbium and uranium are due to vibrations of these neutralizing electrons, and that the forces acting upon these are considerably different from those acting on the ring electrons, which, in many cases, give a normal Zeeman effect. It is probable that these neutralizing electrons play the greatest röle in the optical properties of bodies, such as the properties determining the index of refraction, the extinction coefficient, etc.

Usually the equation of motion of such an electron is given by an equation like the following when a light wave of an electric field $E \cos pt$ is passing by it:

$$m\frac{d^2x}{dt^2} + k\frac{dx}{dt} + n^2x = E\cos pt.$$

where m is the total mass (electromagnetic and material) of the electron, $\kappa \cdot dx/dt$ is the damping or frictional term and n^2x is the quasielastic force. It is an experimental fact as shown by the above work and the work of other investigators, that κ and n^2 are not only functions of the electron and the atom, but that they are also functions of the physical and chemical conditions existing in the neighborhood of the atom.

On the other hand, the effect on κ and n^2 for a ring electron, when external physical and chemical conditions are changed, is very small. It is for this reason, and the probable fact that there are relatively few neutralizing electrons, that we believe that much greater progress can be made in determining some of the properties and constitution of various interatomic systems of atoms and molecules by the study of the absorption spectra of uranium and neodymium than by a study of the arc or spark spectra of the same.

The method of attacking the above problem will be to study the effect on the spectra of a body produced by changing the physical and chemical conditions about the light absorbers or emitters within as wide ranges as possible. Some of the possible changes that

can be made are as follows: Take for instance the uranyl group UO_2. We can find the effect upon the absorption bands produced (1) by diluting the solution, (2) by changing the acid radicle to which the uranyl group is united, (3) by changing the solvent and using mixtures of solvents, (4) by adding other salts (like aluminium chloride), or (5) by adding acids of the same kind, as that of the salt of the uranyl group. The effect of adding foreign salts and acids at the same time and then varying the solvent, or the temperature, can also be tried. In this way a very large number of very interesting things can be tested. In most of these changes *lc* will be kept constant.

In the above examples the temperature (7), the external pressure (8), the electric field (9) and the magnetic field (10) can be changed between wide limits. The latter effect is a very important one. For example, in aqueous solution neodymium salts give a large number of fine bands, in glycerol there are quite a number of new bands replacing the "water" bands, and in the alcohols there are various "alcohol" bands. At low temperatures these bands become very fine and it is quite possible to detect the Zeeman effect. Now it seems quite probable that a "glycerol" band and an "alcohol" band that seem to replace each other as the solvent is changed are both due to the same vibrator. If the Zeeman effect is the same in both cases it would be a strong argument in favor of the above theory. A case that will soon be described is very important. It was found that certain neodymium lines in a pure water solution did not have their wave-length changed when the temperature was changed from 0° to 90°. If, however, calcium chloride was added, then on raising the temperature the above bands were shifted to the red. A very interesting and important investigation is whether the Zeeman effect on this band would be affected by the presence of bodies like calcium chloride.

To be compared with the above changes are changes in the absorption spectra of the crystals of the salt (11) as affected by water of crystallization, or by the presence of foreign substances, or as affected by the polarization (12) or direction of passage of light through the crystal. The absorption spectra (13) of the anhydrous powder at different temperatures, etc., should be found. The

phosphorescent spectrum (14) should be studied in this connection, especially as affected by the mode of stimulation (X-rays, cathode rays, heating or monochromatic light of different wave-lengths). The temperature, electric or magnetic field could be changed about the phosphorescing body. The effect of change of state (15) should be tried if this is possible, also any possible changes of valency of the atoms (16) composing the body investigated. We shall attack the problem from all of these standpoints.

After correlating the data obtained by the above named investigations it is pretty certain that it will be possible to take each vibrator and trace the effects produced upon it by the above changes. It is also quite certain that we shall also know something of the nature of the vibrating system and the part that it plays in that complex body we call the atom. We shall now describe a few results obtained by changing the concentration and temperature of a solution of the chemical compound whose absorption spectrum we are studying.

IX. The Effect of Rise in Temperature on the Absorption Spectrum of Certain Salts in Aqueous Solution.

(*a*) *Uranous Chloride* (*B, Plate VII.*).

To a normal solution of uranyl chloride in water was added a small amount of hydrochloric acid and zinc. The production of hydrogen reduced the uranyl to the uranous state. The same can be done in some cases by simply passing hydrogen gas through the uranyl solution. The solution was placed in the glass trough and a temperature run made as in the usual manner. The thickness of layer was 1 mm. The length of exposure was 50 sec. to the Nernst glower and 4 mins. to the spark, the current through the glower being 0.8 amperes and the slit width 0.20 mm. Starting with the strip nearest the comparison scale the temperatures were 8°, 17°, 33°, 48°, 62° and 73°. An exposure was also made at 80° which is not shown in the spectrogram *B*.

At 8° a mist formed on the prisms and for this reason the spectrum film taken at this temperature is much underexposed and the bands appear wider than at the higher temperatures. At this tem-

perature there is complete absorption of the shorter wave-lengths to λ 3650. A blue-violet absorption extends between λ 4050 and λ 4450. Following this band are three strong bands of about equal intensity and each almost 100 Ångström units wide. Their wave-lengths are approximately λλ 4590, 4760 and 4970. Following is a band at λ 5500, a wide band from λ 6400 to λ 6630 and a rather narrow band at λ 6740.

The absorption does not change very greatly until a temperature of 60° is reached. Above this temperature the increase in absorption is quite rapid as the temperature rises. At 73° the ultra-violet band has widened to λ 3800, the blue-violet band covers the region from λ 4050 to λ 5000. The bands λλ 4600, 4770 and 4980 at 8° have shifted slightly to the red with rise of temperature.

None of the other bands seem to shift to the red at all and the broadening seems to be quite symmetrical. The band at λ 5500 has become about twice as wide as it was at the lower temperatures and the two red bands have merged into one band running from λ 6350 to λ 6800. Between 73° and 80° the absorption increases very greatly. All short wave-lengths are absorbed up to λ 5050. The band in the green runs from λ 5450 to λ 5600 and the band in the red has also widened very greatly, extending from λ 6200 to λ 6800.

(*b*) *Copper Bromide* (*A and B, Plate VIII.*).

The two spectrograms showing the absorption spectra of copper bromide in water for various temperatures were made for different concentrations of the salt. *A* gives the absorption of a 2.06 normal solution 1 mm. thick and *B* the absorption of a 0.25 normal solution 8 mm. thick. The time of exposure to the Nernst glower was 2 mins. (current 0.8 amperes and slit width 0.20 mm.) and to the spark 6 mins. Starting with the strip nearest the comparison scale the temperatures at which exposures were made for *A* were 6°, 17°, 30°, and 45°; for *B* 6°, 17°, 31°, 46°, 59°, 71°, and 85°.

As the spectrograms show, the effect of change of temperature on the absorption spectrum is very marked. Above 45° the concentrated solution did not transmit enough light to affect the photographic film.

(*c*) *Chromium, Calcium and Aluminium Chlorides* (*A and B, Plate IX.*).

A, Plate IX., represents a spectrogram showing the effect of rise of temperature on an aqueous solution of chromium and aluminium chlorides. The concentration of the chromium chloride was 0.125 normal, and of the aluminium chloride 2.28 normal. The depth of layer was 9 mm. The length of exposure to the Nernst glower was 4 mins. (current 0.8 amperes and slit width .20 mm.) and to the spark 6 mins. Starting with the strip adjacent to the comparison scale the temperatures were 6°, 19°, 36°, 51°, 66° and 81°.

The most marked effect of the aluminium chloride was the production of a very pronounced unsymmetrical broadening, which does not occur when a pure aqueous solution of chromium chloride is heated. At 6° the ultra-violet band extends to λ 3000, at 81° to λ 3300, a much greater widening than takes place for a chromium chloride solution in water. At 6° the blue-violet band extends from λ 4100 to λ 4600 and the yellow band from λ 5800 to λ 6200. Not only do the red sides of the blue-violet and yellow bands widen out enormously towards the red, but the short wave-length edges of these bands actually move towards the red. This effect is much more pronounced in the changes of temperature from 51° to 66° and from 66° to 81°. At 81° the blue-violet band extends from λ 4150 to λ 5050 and the yellow band from λ 5900 throughout the remaining portion of the spectrum, as far as the film is sensitive. The fine chromium bands in the red do not appear.

B, Plate IX., is a spectrogram, giving the absorption spectrum of a .125 normal concentration of chromium choride and a 3.45 normal concentration of calcium chloride in water at different temperatures. The length of the solution was 9 mm., the length of the exposures to the Nernst glower were for 5 min. and to the spark for 6 min. The current through the glower was 0.8 amperes and the slit width 0.20 mm. Starting with the strip adjacent to the comparison scale the temperatures at which the exposures were made were 6°, 19°, 31°, 45°, 64° and 80°.

The effect of rise of temperature upon the absorption spectrum of a mixture of chromium chloride and calcium chloride is very similar to the effect on the mixture of chromium chloride and

aluminium chloride. The blue-violet and the yellow bands widen unsymmetrically and the short wave-length edges of these bands apparently moves towards the red at the higher temperatures.

At 6° the ultra-violet band extends to λ 2800, the blue-violet band from λ 4000 to λ 4400 and the yellow band from λ 5600 to λ 6100. At 64° the ultra-violet band extends to λ 3100, the blue-violet band from λ 4000 to λ 4600 and the yellow band from λ 5650 to λ 6300. At 80° the ultra-violet band extends to λ 3250, the blue-violet band from λ 3950 to λ 5000 and the yellow band from λ 5700 throughout the red end of the spectrum as far as the film is sensitive.

(*d*) *Uranyl Chloride* (*A and B, Plate X.*).

A spectrogram (*A*, Plate X.) was made of the absorption spectrum of a normal aqueous solution of uranyl chloride, the depth of cell being 3 mm. Exposures were made to the Nernst glower for 90 sec., current 0.8 amperes and slit width 0.20 mm. The time of exposure to the spark was 6 min. Starting from the comparison spectrum the temperatures were 6°, 18°, 34°, 52°, 68° and 82°.

At 8° the ultra-violet band extended to λ 3550, the blue-violet band from λ 4000 to λ 4450. The bands *a*, *b* and *c* appeared, but the *a* band is very faint. The wave-lengths of the *b* and *c* bands were λλ 4565 and 4725.

At 82° the ultra-violet band extends to λ 3700, and the blue-violet band from λ 3950 to λ 4600. At this temperature only the *b* band appears,—*a* being very weak and *c* being completely merged into the blue-violet absorption band. The *b* band is located at λ 4755.

A spectrogram, *B*, Plate X., was made of a uranyl chloride water solution 0.0156 normal concentration and a depth of layer of 196 mm. Exposures were made to the Nernst gower for 30 sec., current 0.8 amperes and slit width 0.20 mm. No exposures were made to the spark except for comparison spectra. Starting with the numbered scale the temperatures were 6°, 18°, 29°, 44°, 59°, 71° and 79°.

For this concentration there is a very slight temperature effect. There is a very faint transmission band between the ultra-violet and blue-violet bands. This is extremely faint and is practically un-

affected by temperature. The blue-violet band widened slightly with rise in temperature. The uranyl bands in the concentrated solution were much stronger and wider than in the dilute solution.

(*e*) *Neodymium Salts.*

A spectrogram (*A*, Plate XI.) of the absorption spectrum as affected by change of temperature was made of neodymium chloride solution in water, the concentration being 3.4 normal and the depth of layer 12 mm. The length of exposure was 2 min. to the Nernst glower, current 0.8 amperes; slit width .20 mm. The time of exposure to the spark was 6 min. Starting with the strip nearest the numbered scale the temperatures were 11°, 22°, 33°, 45°, 59°, 73° and 85°.

An absorption band appears at about λ 2970 for the 11° temperature, a very strong band from λ 3250 to λ 3285 and an adjacent band from λ 3285 to λ 3310. At 11° a very narrow and feeble transmission band separates these two bands. At 85° the transmission band has weakened very much. At 11° a very strong band lies between λ 3490 and λ 3580. The band λ 4274 is about 8 Ångström units wide. An extremely narrow band appears at λ 4297, λ 4306 and λ 4324. At λ 4234 is a wider and rather diffuse band, it being about 12 Ångström units wide. Bands at 11° lie between λλ 4415 and 4470, λλ 4580 and 4650, λλ 4665 and 4710, λλ 4740 and 4775, λλ 4815 and 4835, and the very wide bands λλ 5010 and 5300 and λλ 5665 and 5935. Weak bands are located at λ 4645, λ 4800, λ 5320, λ 6235, λ 6255, λ 6280, λ 6305 and λ 6380. Rather diffuse bands appear at λλ 6780 and 6840, at λ 6850 and from λ 6870 to λ 6920.

The effect of rise of temperature from 11° to 85° is quite noticeable, although it is not great. In the ultra-violet there is a slight increase in the general absorption. The band λλ 3285 and 3310 widens slightly. The band λλ 3490–3580 at 11° has widened so that at 85° it extends from λ 3450 to λ 3600. The band at λλ 4415 and 4470 has widened but little. The group of bands from λ 4600 to λ 4800 have also widened but little. The faint diffuse bands λλ 4645 and 4800 have practically disappeared. The bands λλ 5010 and 5300 and λλ 5665 and 5935 at 11° have widened at 85° to

λλ 5010 and 5350 and λλ 5660 and 5985. The widening of the latter band is distinctly unsymmetrical. The existence of the band λ 5320 causes the band λ 5010 to λ 5300 to widen unsymmetrically.

The bands in the region λ 6300 become less sharp as the temperature rises. At 11° there was considerable transmission in the region λ 6850. At 85°, however, this transmission disappears and there is practically complete absorption from λ 6760 to λ 6920. The very sharp bands λλ 4282, 4300, 4310, 4322 and 4343 do not appear to change very much with change in temperature. On the strip taken at 73° these bands appear sharper than on any of the other strips.

A spectrogram (*B*, Plate XI.) showing the effect of rise in temperature was made on a neodymium chloride solution in water of c.17 normal concentration and a depth of layer of 196 mm. The amount of neodymium chloride in the path of the light is approximately the same as in the spectrogram, showing the effect of temperature upon a 3.4 normal concentration and a depth of cell of 12 mm. In this case the temperatures were 5°, 16°, 28°, 42°, 59°, 72° and 82°. Exposures were made to the Nernst glower for 3 mm. current 0.8 amperes and slit width 0.20 mm. Each strip was exposed to the spark for 6 mm. The purpose of making this spectrogram was to find the effect of concentration of a salt upon the changes produced by change in temperature.

A description of the bands at 5° and 82° will be given. Any change between these two temperatures that takes place is a gradual one. Transmission begins at λ 2600. Bands appear between λλ 3250 and 3300 and λλ 3455 and 3575. The band λ 4274 is much sharper and narrower than for the more concentrated solution. The numerous fine bands in the region λ 4300 are very faint. The bands λλ 4420 to 4460, λλ 4600 to 4630, λ 4645, λλ 4680 to 4705, λλ 4745 to 4770 and λ 4820 have rather diffuse edges. Wide bands appear from λ 5020 to λ 5290 and from λ 5685 to λ 5920. Diffuse bands are located at λ 5310, λ 6810 and λ 6900. The group in the region λ 6300 appear, but they are extremely faint.

At 82° the general absorption has increased in the ultra-violet and has reached to about λ 2800. It will be noticed here that the effect of rise in temperature is greater upon this general ultra-violet

absorption in the dilute solution, than it is for the concentrated solution previously described.

The band λλ 3455 to 3575 at 5° has widened slightly, having the limits λλ 3445 and 3580 at 82°, the widening being about 15 Ångström units. This band in the concentrated solution widened 60 Ångström units. Practically no effect of temperature is to be noticed upon the bands from λ 4200 to λ 4900 with rise in temperature. At the higher temperatures the bands are slightly more diffuse, but this change is very small. The band λλ 5020 to 5290 at 5° has widened to λλ 5015 and 5285, about 10 Ångström units. The corresponding widening for the concentrated solution was approximately 50 Ångström units, although it must be noted that in the more concentrated solution this widening was mostly due to the increased absorption of the band λ 5310 at the higher temperatures. The band λ 5685 to λ 5920 at 5° has widened to λλ 5775 and 5930, about 20 Ångström units, compared with a widening of 55 Ångström units for the more concentrated solutions. None of the other bands show any appreciable change with change in temperature.

A spectrogram (*A*, Plate XII.) was made showing the effect of temperature upon the absorption spectrum of a 1.66 normal aqueous solution of neodymium bromide, the depth of layer being 6 mm. An exposure of 4 mm. was made to the Nernst glower, at .8 amperes and a slit width of 0.20 mm. The length of exposure to the spark was 6 mins. The temperatures of exposure, starting with the strip adjacent to the comparison spark, were 4°, 20°, 36°, 50°, 68° and 83°.

At 4° there is complete absorption in the ultra-violet up to λ 2600. A broad absorption band appears at λ 2660 to λ 2800 and from λ 2950 to λ 3060. These absorption bands appear with a more or less general absorption. Bands appear at λλ 3460, 3500 and 3540. The band at 4274 is weak. Weak and diffuse bands occur at λλ 4440, 4630, 4695, 4760, 4825, 5095, 5260, 6810 and 6900. Wider bands are located at λλ 5116 to 5140, λλ 5200 to 5240 and λλ 5710 to 5850.

At 83° the spectrum is almost exactly the same as at 4°. The ultra-violet absorption is complete up to λ 3050. The bands at λ 3500 have increased in width slightly and the band λ 4274 is

slightly broader. The bands that have widened appreciably are λλ 5195 to 5260 and λλ 5700 to 5880. The change in the absorption is greatest when the temperature is changed from 68° to 83°. Up to 68° there is practically no change in the absorption spectrum at all.

A spectrogram (*B*, Plate XII.) showing the effect of temperature was made, using an aqueous solution of .055 normal concentration of neodymium bromide, the depth of the layer being 197.4 mm. This spectrogram was made to compare with that taken with a 1.66 normal concentration of the same salt and a depth of layer of 6 mm. The exposures to the Nernst glower lasted 90 sec. in this case, current 0.8 amperes and slit width of 0.20 mm. The length of exposure to the spark was 6 mins. Starting with the strip nearest to the comparison scale the temperatures of the solution were 5°, 16°, 29°, 42°, 55°, 68° and 84°.

At 5° there is practically complete transmission of light between λ 3400 and λ 2600, no ultra-violet bands appearing, as was the case for the more concentrated solution. The bands λλ 4445, 4693, 4760, 4825 and 5095 were somewhat sharper than they were in the concentrated solutions. The two largest bands extended from λ 5200 to λ 5250 and from λ 5710 to 5850. As in the case of the more concentrated solution, so here, the greatest change in the absorption took place in the change from 68° to 84°. The ultra-violet absorption increased up to λ 2900. The bands at λ 3500 became considerably stronger, but they widened very little. The bands λλ 4445, 4693, 4760 and 4825 are somewhat weaker than at 5°. The wide bands remained practically as wide as at 5°, λ 5200 to λ 5250 and λ 5705 to 5870. This indicates a widening of about 25 Ångström units for the latter band. For the more concentrated solution the widening of these two bands was 25 and 40 Ångström units respectively. It is thus seen that in the more concentrated solutions the bands widen more with rise in temperature than they do in the less concentrated solutions. At 42° in the dilute solution there appears a narrow band at λ 6710. This increases in intensity with rise in temperature. This band does not appear at all in the concentrated solution.

A spectrogram (*A*, Plate XIII.) was made of neodymium chloride and calcium chloride in water. Exposures were made for 30 sec.

to the Nernst glower, the current being 0.8 amperes and the slit width 0.20 mm. The length of exposure to the spark was 4 mins. Starting with the strip nearest the numbered scale the temperatures were 6°, 17°, 31°, 49°, 63°, 74° and 82°.

The general effect of the addition of calcium chloride is to make all the bands hazier, and to increase the transmission throughout the region of the band. At 6° there is a slight transmission throughout the ultra-violet portion of the spectrum. As the temperature is raised this general transmission is decreased, and at 82° practically no light passes through the solution of shorter wave-length than λ 2800. Sharp bands occur at λ 3464, λ 3500, λ 3535, λ 4276 and weak diffuse bands at λ 4295, λ 4305, λ 4340, λ 4445, λ 4620, λ 4695, λ 4760, λ 4825, λ 5095, λ 5130, λ 5225, λ 5260, λ 5320, λ 5710, to λ 5860, λ 6245, λ 6810 and λ 6900.

At 82° the bands in the λ 3500 region are slightly more intense than at 6°. Practically all the bands from λ 4200 to λ 5200 have become much weaker at the higher temperature. This is especially true of the band λ 4276, its intensity being less than half what it is at 6°. Most of the bands are shifted to the red with reference to the same bands at 6°. For instance, λ 5095 is shifted 5 Ångström units towards the red. The bands λ 4695, λ 4760 and λ 4825 are all shifted to the red at the higher temperature, and especially λ 4825, the shift in thise case amounting to 5 Ångström units. In the case of these bands the shift is not an apparent one due to unsymmetrical broadening, for in this instance there is no broadening at all.

The band from λ 5710 to λ 5860 at 6° has widened very unsymmetrically and has the limits λ 5710 to λ 5920. The short wave-length side is quite sharp and its position is practically independent of the temperature. The long wave-length edge is quite broad and recedes quite rapidly towards the red as the temperature is raised. The bands in the red λλ 6810 and 6900 grow fainter and fainter with rise in temperature, and have practically disappeared at 82°. The band λ 6245 is very weak at 6° and has disappeared at about 60°.

It will thus be seen that not only does the presence of calcium chloride modify greatly the absorption of neodymium chloride, but that it changes the effects due to temperature very fundamentally. In pure neodymium chloride practically no bands decrease in in-

tensity with rise in temperature, and at present no shift has been detected. When calcium chloride is added to the solution most of the bands decrease in intensity with rise in temperature and several are shifted towards the red at the same time. Several bands disappear. Moreover, the band λλ 6800 to 6900, although it widens, this widening is entirely on the red side, whereas for the pure neodymium chloride solution this widening always takes place on both sides of the band.

A spectrogram (*B*, Plate XIII.) was made to show the effect of change in temperature upon a 2.15 normal aqueous solution of neodymium nitrate. The length of layer was 3 mm. The exposures were for 40 sec. to the Nernst glower, current 0.8 amperes, slit width .20 mm. The length of exposure to the spark was 6 mins. Starting with the strip nearest the comparison spark the temperatures were 4°, 17°, 29°, 43°, 58°, 71° and 84°.

The changes in the spectrum due to this change in temperature of 80° was very slight. The NO_3 band extends to about λ 3250 at 4°, and to about λ 3280 at 84°. The bands at λ 3500 became considerably wider and their edges more diffuse at the higher temperatures. At the lower temperatures fine bands appear at λλ 5210, 5225 and 5240. At 84° these bands all merge into a single band. The red band extends from λ 5705 to 5860 at 4°. The band at λ 5820 is very faint at the lower temperatures. At 84° it is unrecognizable, At this temperature the red band extends from λ 5700 to λ 5880. The widening of this band for the concentrated solution is somewhat greater than for the dilute solution, but the effect of concentration is very slight. This is to be expected since the effect of temperature itself is so very minute.

A spectrogram (*A*, Plate XIV.) was made of an aqueous solution of neodymium bromide 1.66 normal concentration and 54.6 mm. depth of cell. The exposures were 3 mins. to the Nernst glower and 6 mins. to the spark. The current in the Nernst glower was 0.8 amperes and the slit width 0.20 mm. Starting with the strip nearest the comparison scale the temperatures were 6°, 20°, 33°, 47°, 62°, 73° and 82°.

The effect of rise in temperature upon the absorption spectra of this salt was quite marked; practically all of the bands broaden-

ing and becoming more intense. At 6° the ultra-violet absorption extended to λ 3600. At 82° it had advanced to λ 3800. Very narrow and fine bands appear at λ 4186, λ 4300, λ 4308, 4345, 6240, 6265, 6290, 6305, and much broader bands at λ 6380 and λ 6740. Wide bands occur at from λλ 4390 to 4480, λλ 4550 to 4850, λλ 4990 to 5340, λλ 5650 to 5950 and λλ 6760 to 6930, at 6°. At 82° these bands have the following limits respectively: λλ 4380 to 4500, λλ 4540 to 4910, λλ 4960 to 5370, λλ 5620 to 5990 and λλ 6730 to 6960.

(*f*) *Erbium Chloride.*

A spectrogram (*B*, Plate XIV.) was made to show the effect of rise in temperature upon the absorption spectrum of a solution of erbium chloride. For this purpose a 0.94 normal solution of erbium was used and the depth of layer was 48 mm. The solution probably contained a considerable number of impurities, so that in fact the amount of erbium was quite small. The absorption spectrum was found to change but little with rise in temperature, thus indicating a dilute solution. Exposures were made for 30 sec. to the Nernst glower and 4 mins. to the spark. The current through the glower was 0.8 amperes and the slit width 0.20 mm. Starting with the spectrum nearest the comparison scale the temperatures were 7°, 17°, 29°, 46°, 60°, 70° and 80°.

At 70° the ultra-violet is absorbed to λ 3950. As the temperature is raised the ultra-violet absorption increases, and at 80° it reaches λ 3150. Bands from 20 to 40 Ångström units wide occur at λ 3235, λ 3510, λ 3640 and λ 3785. These bands are slightly wider at 80°, but as for all the other erbium bands this widening is very small. Weak and narrow bands appear at λλ 4165, 4425, 4458, 4500 (strong), 4535, 4540, 4555, 4580, 4685, 4750 (30 Å. u. wide), 4810, 4840, 4855, 4870 (strong and 20 Å. u. wide), and 4920, λ 4920 lies alongside of a fuzzy band extending from λ 4910 to λ 4950.

After these comes a rather wide band which for a shorter length of layer would most likely be broken up into a number of much finer bands. This band extends from λ 5190 to λ 5250. At λ 5217 there runs a narrow sharp line through the fuzzier and wider band. Broad (about 30 Å. u. wide) and very faint bands are located at λ 5630 and λ 5760. For greater concentrations these would prob-

ably show as finer bands. The band at λ6540 is much more diffuse on the red than on the violet side; this possibly being due to a component that is not separated at this temperature. Other bands are located at λλ 5365, 5380, 5425, 5445, 5505, 6410, 6440, 6495 and 6690.

The general effect of rise in temperature here is to cause the lines to become slightly fuzzier and to show more of a "washed out" appearance. No shift due to rise in temperature was noticed.

Throughout all the previous work the wave-lengths were read directly from a scale. This scale was made so as to give the wave-lengths in Ångström units directly. It was found in the measurements that the Seed films did not correspond to the Wratten and Wainwright films, when the same spark spectra on the two kinds of films were placed beside one another. This was probably due to different shrinkage of the two kinds of films on fixing, washing and drying. For this reason the wave-length measurements are not intended to be absolutely correct but only relatively so. All the temperature work was done with Wratten and Wainwright films. The relative measurements of fine bands for any spectrogram are probably correct to within a few Ångström units.

X. Summary.

The absorption spectra of the uranyl salts contain a series of bands in the blue and violet. Twelve of these bands can usually be detected for each salt. Starting from the blue end of the series the bands are designated by the letters *a*, *b*, *c*, etc. These bands are usually diffuse and from 30 to 50 Ångström units wide.

The uranyl bands of uranyl nitrate in water are all farther to the violet than the uranyl bands of any other salt investigated, or of uranyl nitrate in other solvents.

The uranyl absorption bands of crystals of uranyl nitrate agree with the absorption bands of an aqueous solution of the nitrate, with the exception of the *f*, *g*, *h* and *i* bands; these latter being shifted to the red in the crystal.

Dilution within the ranges studied does not affect the position of the uranyl bands. Theoretically, all the uranyl salts in water

should give the bands of the same wave-lengths for very dilute solutions.

The uranyl bands of the nitrate in methyl alcohol are all shifted to the red about 50 Ångström units, with reference to the bands in water. Mixtures of water and methyl alcohol show that we have both sets of bands existing for the same solution, the "water" bands increasing in intensity as the amount of water increases. The water bands are the more persistent. This indicates the existence of a hydrate and an alcoholate of the uranyl group. In ethyl alcohol the *a*, *b*, *c* and *d* bands are shifted to the red with reference to the methyl alcohol bands. The other bands appear to have the same positions a. the methyl alcohol bands.

The absorption spectrum of the anhydrous salt is very complex and the bands could not be recognized.

The bands of uranyl bromide in water, of uranyl acetate in water and methyl alcohol, and also of the anhydrous salt, are approximately of the same wave-lengths, differing but slightly from the wave-lengths of the uranyl nitrate bands of an aqueous solution.

The bands of uranyl sulphate in water are all shifted towards the red about 50 Ångström units, with reference to the uranyl nitrate bands in water. For both the sulphate and nitrate in water the bands are very much alike. The *i* band is very weak in both cases.

Uranyl chloride bands of an aqueous solution are shifted to the red with reference to the uranyl nitrate bands of an alcoholic solution. The addition of calcium chloride or aluminium chloride is found to produce very marked effects upon the uranyl chloride bands. The addition of sufficient aluminium chloride to a water solution of uranyl chloride, or of calcium chloride to a methyl alcohol solution of uranyl chloride is found to cause the *d* and *e* bands to come together, so as to form a single wide band, and to cause the other uranyl bands to shift so that the whole resulting series of bands is almost identical with the series of bands of an ethyl alcohol solution of uranyl chloride. The effect of adding foreign substances also greatly modifies the intensity of the bands. An example of this difference of action is the effect of adding aluminium chloride to an aqueous solution of uranyl chloride. The

a and *b* bands are affected entirely differently; the *a* band being very much reduced in intensity and made narrower, whereas the *b* band becomes very much stronger and wider.

A new set of fine bands in the green has been discovered in the absorption spectrum of an aqueous solution of uranyl chloride. These only appear for pure water solutions; a small amount of aluminium or calcium chloride causing them to vanish. They do not appear for methyl or ethyl alcohol solutions, and for no other uranyl salt except very faintly for the sulphate.

The absorption spectrum of several uranous salts has been photographed. The spectrum is entirely different from that of the uranyl compounds. The absorption spectra of uranous chloride in methyl alcohol and in water were found to be very different. The absorption spectrum of neodymium chloride in glycerol was found to be entirely different from that of the salt in water. Mixtures of water and glycerol seem to indicate the existence of both sets of bands in the same solution. The "glycerol" bands are more persistent with reference to water bands than "alcohol" bands are. Much more work along this line is contemplated.

Rise in temperature has been found in general to cause an increase in the amount of absorption, and to cause the absorption bands to widen. This widening of the bands may or may not be symmetrical.

Some of the absorption bands of uranous chloride widen very rapidly with rise in temperature. Other bands do not widen so rapidly, and seem to be slightly shifted towards the red.

In solutions containing a single salt, it has invariably been found that the bands widen with rise in temperature, and that this widening is greater, the greater the concentration of the solution.

The uranyl bands of aqueous solutions of the chloride and sulphate of uranyl are shifted towards the red with rise in temperature. The intensity of the uranyl bands does not seem greatly modified by changes in temperature.

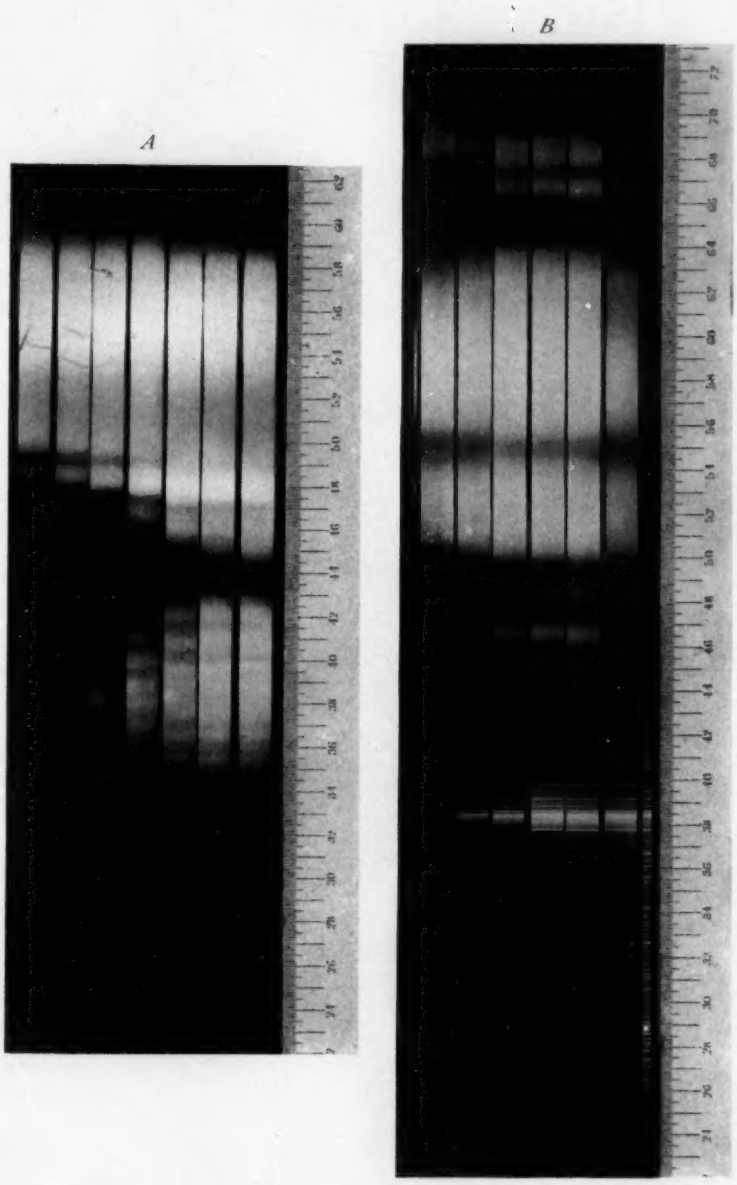
The effect of rise in temperature on the absorption syectrum of a solution of a salt containing calcium or aluminium chloride is very peculiar. The bands usually broaden very unsymmetrically,

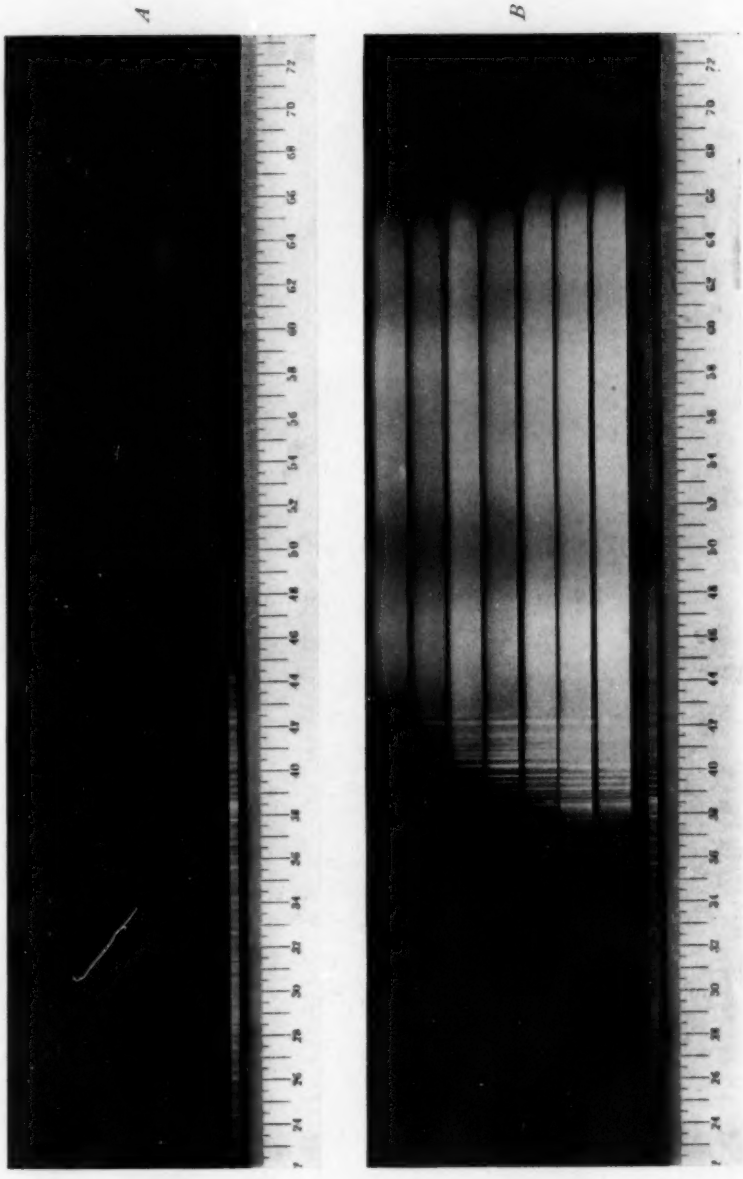
and in all cases investigated, the widening has been on the longer wave-length edge. A tyipcal example is shown in Plate III.

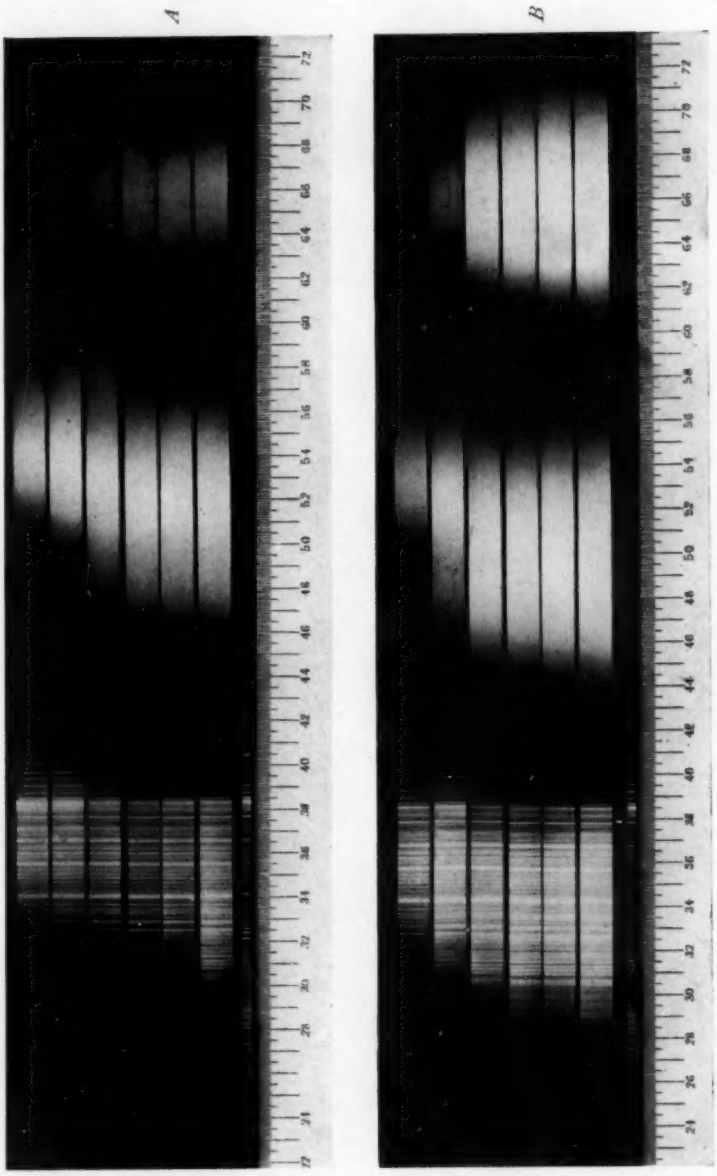
Rise in temperature causes the neodymium bands to widen slightly, but no shift of the bands has been noticed. However, when calcium chloride has been added to the neodymium solution, a rise of temperature causes many of the bands to become much less intense, and also causes some of the bands to shift to the red. In the recent work of Becquerel and others it is quite possible that the presence of various foreign bodies in the crystals along with the neodymium may have a very great influence upon the absorption spectrum.

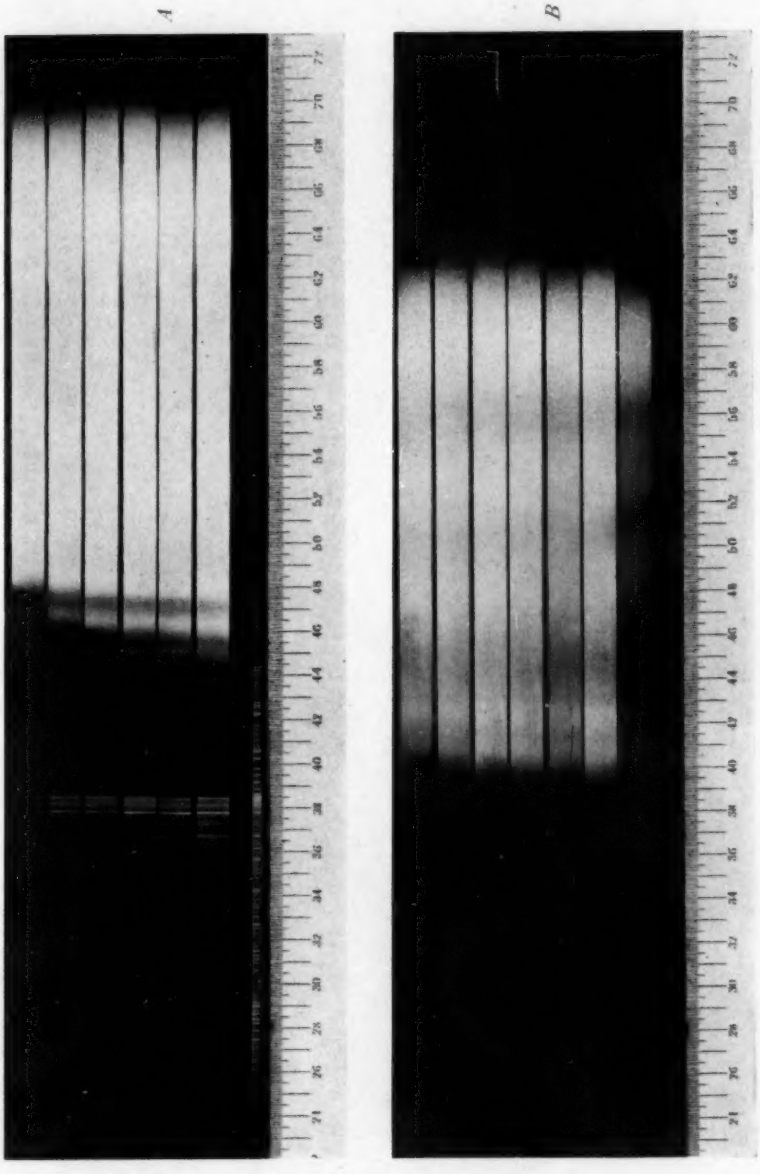
All the above conclusions must be understood to be limited to the conditions and within the ranges described in the earlier parts of this paper.

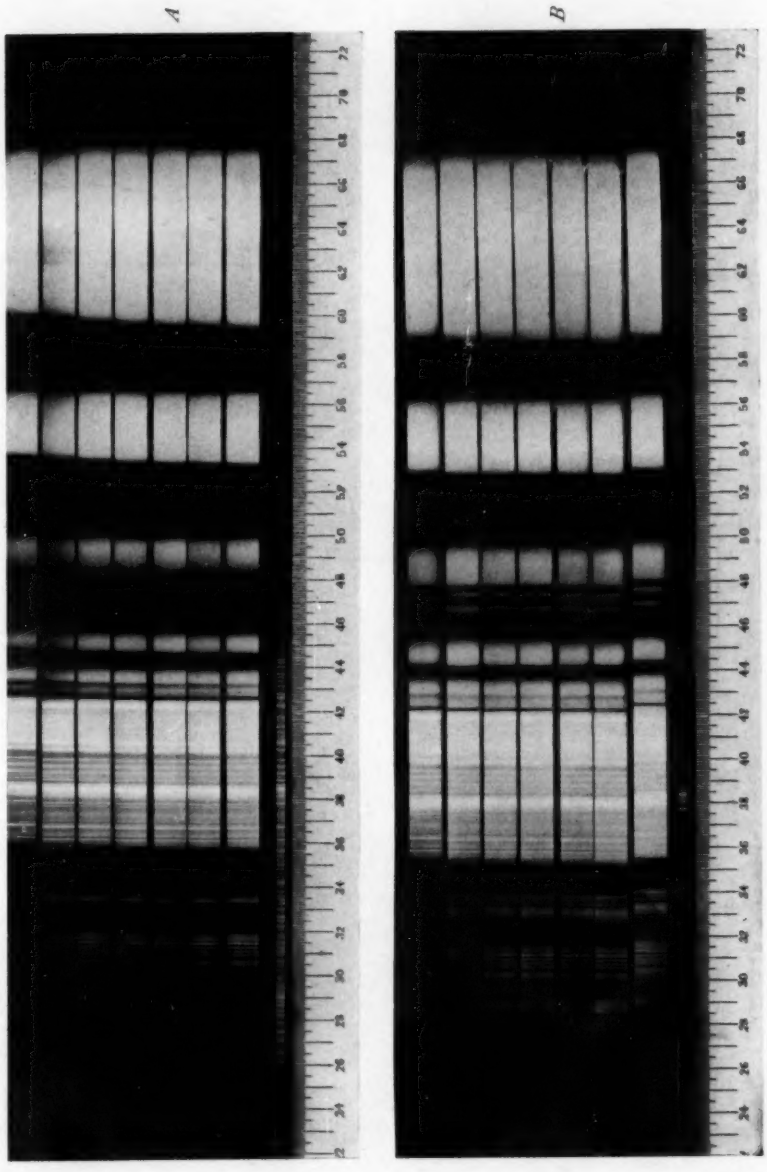
PHYSICAL CHEMICAL LABORATORY,
JOHNS HOPKINS UNIVERSITY,
May, 1909.

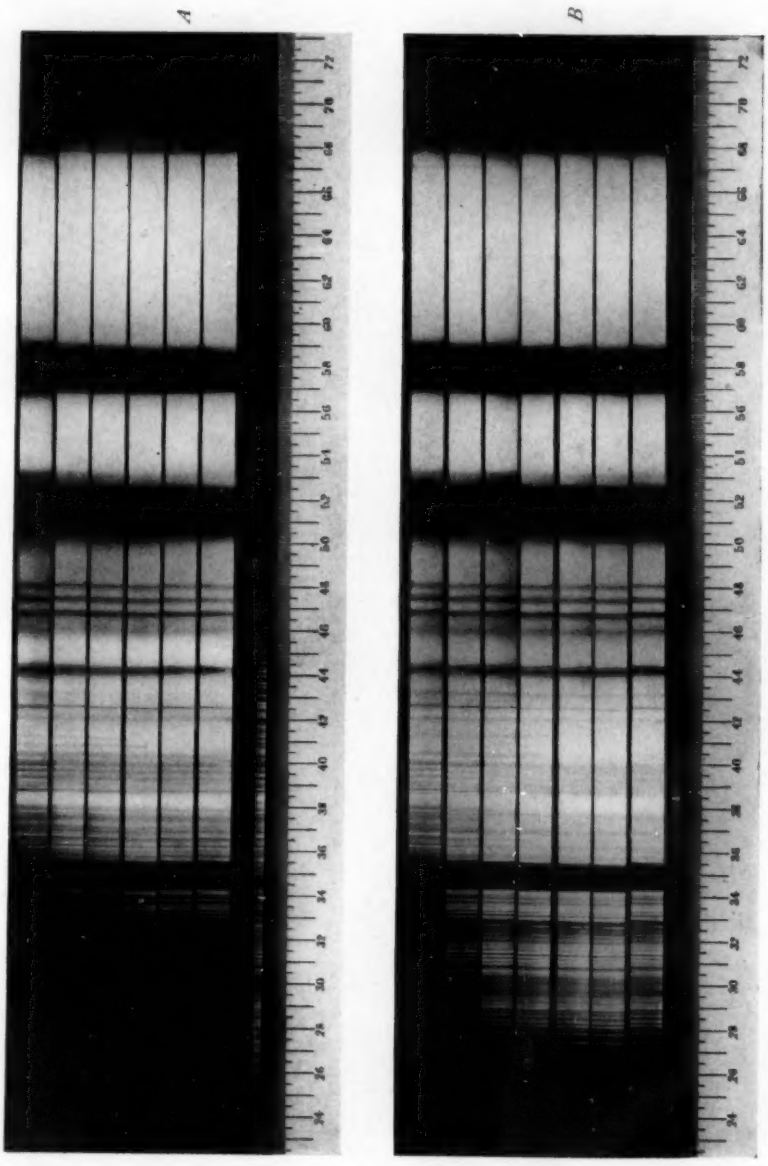

A
B

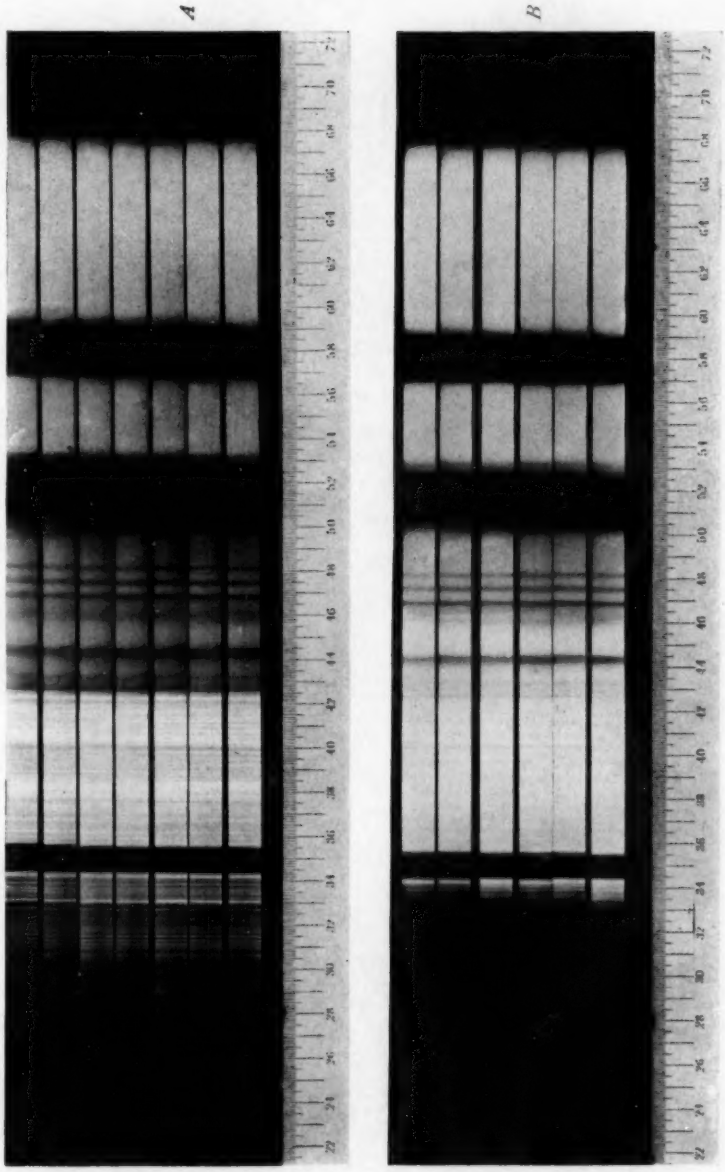

A
B

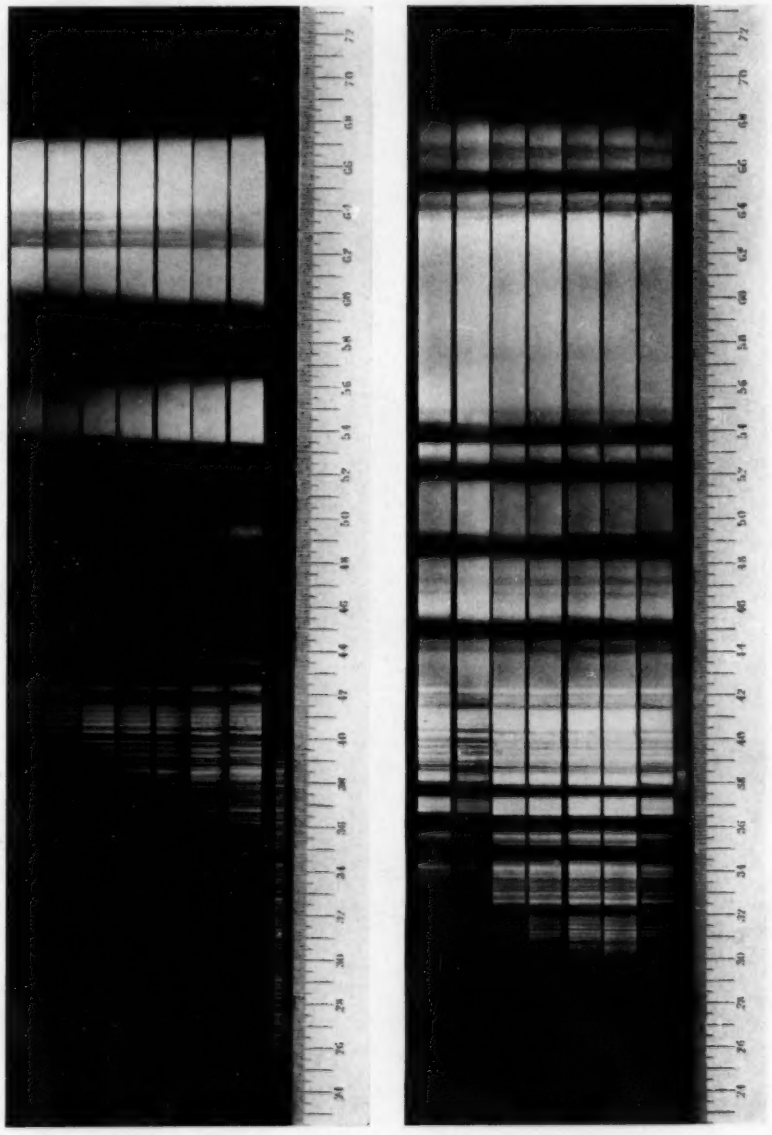

A
B


A
B


A
B


A
B


A
B



EARTHQUAKES: THEIR CAUSES AND EFFECTS.

By EDMUND OTIS HOVEY.

(*Read April 24, 1909.*)

The occurrence of three earthquakes in the western hemisphere within the space of nine months in 1906–1907, all of which were attended with disastrous effects upon human life and property, attracted as never before the attention of the world, and particularly of the United States, and focused interest upon the science of seismology in a manner calculated to advance materially the study of movements and other physical changes in the earth's crust. San Francisco, in April, 1906, Valparaiso, in August, 1906, and Kingston, in January 1907, attracted wide notice, but the disaster that overwhelmed Messina, Reggio and vicinity on December 28, 1908, capped the climax, and sufficient reason is apparent for the universal interest now prevailing, one manifestation of which is the present symposium. The thesis of the seismologists is that the chain of earthquake observatories that have been established in the past decade and a half should be extended and united into a network of stations covering the globe, sufficiently, at least, to furnish a complete record of the important vibrations propagated through the earth, indicate their places of origin and provide data for more satisfactory theories as to their causes.

Great earthquakes rank with volcanic eruptions as being the most terrifying of all natural phenomena. Usually coming with no recognized warning, often happening in the night, extremely indefinite as to source, extent and duration, they fill the mind of the human observer with the horror of utter helplessness. They have been far more destructive to human life and property than volcanic eruptions have been, for we have the earthquake shocks of Sicily, 1693, with 60,000 victims; Yeddo, Japan, 1703 (200,000); Peking, 1731 (100,000); Lisbon, 1755 (60,000); Calabria, 1783 (60,000) and Messina-Reggio, 1908 (200,000); besides many others, to

compare with the volcanic outbursts of Krakatoa, 1883, destroying 36,500 victims; Vesuvius, 1663 (18,000); Mt. Pelé, 1902 (29,000) and the Soufrière of St. Vincent, 1902 (1,400), other historic eruptions having entailed comparatively small loss of life.

Although earthquakes have been recorded frequently throughout all historic time, seismology is one of the youngest of the sciences —it is still in its formative state. Scientific interest in the subject has indeed not been lacking, but real edvance was retarded by the fact that, up to the latter part of the nineteenth century, the causes of the phenomena were sought without rather than within the earth itself. Geology was not seriously called upon for aid in solving the problems.

The modern science of seismology is generally held to have had its beginning with the publication, in 1862, of Robert Mallet's great book upon the so-called Neapolitan, or better Basilicata, earthquake of 1857. Mallet, however, approached his task with the preconceived idea that earthquakes were always caused by subterranean explosions, and his observations and deductions were warped accordinigly. The science received its real start from Eduard Suess, when he published in 1874[1] his brilliant generalization showing the intimate association of more than forty Austrian earthquakes with the already well-known Kamp, Thermen and Mürz fault lines near Vienna and postulated crustal movements as an important cause of seismic disturbances, thus combatting the "centrum" theory of Mallet and others. Suess followed this paper with a still more important paper[2] the next year along the same lines showing the intimate relation of the great earthquakes of southern Italy and Sicily to the fault zones of the region. Impetus was added by the publication of the illuminating treatise of Rudolph Hoernes[3] in 1878, in which earthquakes were first definitely classified into (1) those due to the collapse of the roofs of cavities within the earth's crust, (2) those resulting from explosions connected with volcanic eruptions and (3) tectonic quakes, or those caused by crustal movements along fault planes or due to other effects of the action of mountain-

[1] "Die Erdbeben Nieder-Oesterreichs," *Denkschr. k. Akad. Wiss.*, Wien, XXXIII., Abth. I., p. 61, 1874.

[2] "Die Erdbeben des südlichen Italien," *id.*, XXXIV., Abth. I., p. 1, 1875.

[3] *Jahrbuch d. k. k. Geol. Reichsanstalt*, XXVIII., p. 387, Wien, 1878.

building forces. Many others in Europe, Japan and America have contributed to the advance of seismology, but particular mention should be made of the services of Professor John Milne, of England, whose long residence in Japan and intimate study of the earthquake phenomena of that and other uneasy regions have enabled him to contribute more than any other one person to the advance of the new science.

The perfecting of instruments for the purpose of recording movements of every kind in the surface of the earth has vastly extended our knowledge of the character of earth vibrations and enhanced the value of deductions affecting the theory of earthquakes. The instrumental study of earthquakes by means of seismographs, however, can hardly be said to antedate the year 1892, but within the past decade and a half the number of fully equipped earthquake stations has vastly increased, the growth having been considerably accelerated through the interest aroused by the disasters of the last three years. There are now in Great Britain and her colonies fifty seismographic stations equipped with the same type of instrument, while in all the world there are more than two hundred stations equipped with instruments capable of recording world-shaking earthquakes. More than half of these stations are in Europe.

No large part of the surface of the globe seems to be entirely stable, but certain regions or zones are much more liable than others to the occurrence of earthquakes. If we study a map of the world upon which their location has been plotted, we find in the eastern hemisphere a broad belt of seismic activity extending from west to east through the Mediterranean Sea, Persia, the southern Himalayas and the Sumatra-Java group of islands. A branch zone stretches from the southern end of the Caspian Sea northeastward half way across Asia. This is de Montessus de Ballore's "Alps-Caucasus-Himalayas" belt and it has furnished more than 53 per cent. of recorded shocks.[4] A seismic belt practically encircles the Pacific Ocean, the principal points in it being the Japanese Archipelago, Alaska, California, Southern Mexico and Central America and the

[4] F. de Montessus de Ballore, "Les tremblements de terre," p. 24, Paris, 1906.

northern and southern Andes. This "Circum-Pacific" or "Andes-Japan-Malay" belt has given 41 per cent. of the quakes. In the western hemisphere in addition to a part of the circum-Pacific belt, there are the West India Islands and the mountains of Venezuela forming a seismic zone. Earthquakes mostly of volcanic origin have visited many of the islands of the South Seas. The major portions of Africa and South America remain blank upon such a map, probably because little is known about their seismicity.

We are in the habit of thinking of eastern North America as a region free from earthquake shocks. The impression, however, is erroneous, since New England has experienced about 250 recorded shocks since the Pilgrims landed at Plymouth, and there have been at least four great earthquakes in the eastern half of the continent within the past two and one half centuries, one on the fifth of February, 1663, which affected the St. Lawrence Valley over an area more than six hundred miles long and three hundred miles wide as described in the "Jesuit Relations."[5] In 1811–1812 heavy quakes occurred in the central part of the Mississippi Valley, accompanied with considerable subsidence fifty miles south of the junction of the Ohio and Mississippi Rivers. Strong shocks continued for more than a year and evidence of the sinking still persists in lakes and submerged trees. The southeastern part of the United States was the center of an earthquake shock January 4, 1843, the waves of which were felt at points at least eight hundred miles apart.[6] In 1886 occurred the Charleston earthquake, an event still fresh in the minds of most of our population.

As to earthquakes of the several classes, the falling in of the roof of a buried cavity causes slight shocks. Quakes of this kind have often been reported from certain parts of Switzerland, the Tyrol and elsewhere, but all have been local in character. It seems certain too that the blocks falling in the caverns of southern Indiana and Kentucky produced vibrations sensible on the surface, but reports of such have not come under my eye.

Earthquakes arising from volcanic explosions or associated with eruptions form a much more important subdivision. Until within

[5] W. H. Hobbs, "Earthquakes," p. 315, New York, 1908.
[6] H. D. Rogers, *Am. Jour. Sci.*, I., XLV., 342, 1843.

thirty-five years, indeed, it was the general belief that volcanic earthquakes were by far the most numerous and destructive of all. This idea controlled and vitiated Mallet's work, but it is now known to be erroneous, for although it is true that earthquake zones coincide in part with belts of volcanic activity, shocks are more frequent and more severe in non-volcanic regions. The severest quakes of South America have not happened around the great volcanoes; the shocks of California are evidently independent of the now extinct or at any rate dormant volcanoes of the Cascade Range; the recent (1899) great earthquakes of Alaska were in the vicinity of Yakutat Bay, at a long distance from the active vents of the Aleutian Islands or any recent volcanicity; the earthquakes of Japan are most numerous and severe in the non-volcanic parts of the islands; the great disasters of the Caribbean Sea have occurred in Jamaica and at Caracas, hundreds of miles from Mt. Pelé and St. Vincent's Soufrière, and have not been contemporaneous with any eruptions.

On the other hand, some of the most violent of historic volcanic eruptions have been unattended by severe earthquakes or have given rise to shocks of merely local significance. The Island of Martinique in the French West Indies lies within a markedly seismic zone, but the great eruptive activity of 1902–1903 was free from earthquake shocks. This fact is of particular interest, because the eruptions were of the most highly explosive character. Although, however, no vibrations were felt upon the island of Martinique and no subterranean noises were heard there, dull sounds like the booming of distant cannon were heard the morning of the great eruption of May 8, 1902, at Caracas, Venezuela, 450 miles distant, southwest, where people feared that a naval battle was in progress off their coast. Similar booming was reported from St. Kitts, 200 miles northwest of Martinique and from other regions. I myself was on the island of St. Vincent, 100 miles due south of Pelé when the great eruption of June 6, 1902, occurred, and I felt several dull thuds, as if some heavy object had fallen in a neighboring room. The noises seemed to come from beneath the ground, and they were due, in all probability, to subterranean explosions or to the rushing of lava into underground cavities, somewhat on the priniciple perhaps of the water hammer. On the island of St. Vincent

some observers, indeed, had noted an increase of seismic shocks for a year or more before the volcano burst into violent eruption in May, 1902. The eruption itself, however, was free from earthquakes, except apparently for the quivering of the mountain due to the uprush of steam and ejecta through the conduit, just as happens in the chimney of a fire engine under full blast. The chattering vibrations thus set up in the volcano shook a narrow strip of recent beach formation from the west base of the mountain, where the declivity of the shore is considerable.

Vesuvius being the volcano that has been most continually and thoroughly under observation throughout its known history, we naturally look to its records for light upon the relation between volcanic eruptions and earthquakes. When this old center, which was not known to the ancients as a volcano, renewed its activity in the year 79, the first phase was a series of earth shocks which increased in frequency and severity until the afternoon of August 24, when the eruption actually began. The ground is said to have rocked to and fro like the sea, but we read of no great damage as resulting therefrom even in Pompeii and Herculaneum at the very base of the mountain. The outbreak of 1631 occurred after centuries of repose and was heralded by a half year of earthquakes and terrific noises in the interior of the mountain. This history has been repeated again and again in greater or less degree, particularly when the eruptions have been of the explosive kind. According to the report of A. Lacroix, violent earth movements shook the cone of Vesuvius during the great eruption of April, 1906, and were felt throughout much of the surrounding region. Whatever effects have been produced have been local in extent and comparatively light in degree.

The eruptions of Etna usually have been accompanied by the formation of great fissures in the upper part of the cone, and the opening of these fissures has been accompanied by severe vibrations of the surface of the mountain, as has been vividly portrayed by Silvestri in his account of the eruption of 1879, but the shocks seldom affect the mainland of Calabria across the narrow Strait of Messina. Stromboli, the "Lighthouse of the Mediterranean,"

often shakes its island, but the disturbances are rarely felt in nearby Sicily.

The most violent of all recorded volcanic explosions is that which took place in the Strait of Sunda, August 26–27, 1883, when the volcano of Krakatoa was blown to pieces. This outburst destroyed half the mountain and left soundings of 160 fathoms where part of the cone had formerly stood. It produced sea waves that affected tide gauges half way around the world; air waves that traveled three times around the globe before they ceased to be distinguishable; and it threw dust into the air to such a height that it remained suspended for months, but the earthquake shocks produced were strictly local in character and were scarcely felt at Batavia, 90 miles from the crater.

Another of the great explosions of modern times was that of July 15, 1888, when the Japanese volcano Bandai-san, extinct for a thousand years, burst into sudden eruption. In the immediate vicinity of the mountain a moderately severe earthquake shock lasting about twenty seconds was felt at half past seven in the morning. This was soon followed by additional shocks which culminated when the explosion occurred at the surface, but none was felt severely beyond a limited area.

Even the eruptions of the Hawaiian volcanoes, Kilauea and Mauna Loa, which are the types of the class of "quiet volcanoes," have sometimes been accompanied by severe local earthquakes. Many eruptions of Mauna Loa, indeed, have been recorded of which the first indication to the inhabitants of the town of Hilo only a few miles away has been the light seen at night reflected in the clouds from the streams of flowing lava. On March 27, 1868, however, there began a series of earthquakes on the southern flanks of the mountain which increased in frequency and intensity for a week and culminated in one of the most severe eruptions known in the history of the volcano, during which a great fissure opened, discharging vast quantities of lava that flowed to the sea.

In the words of Dr. Titus Coan,[7] who was on the island at the time:

[7] *Am. Jour. Sci.*, II., XLVI., 107, July, 1868.

Meanwhile the whole island trembled and shook. Day and night the throbbing and quaking were nearly continuous. No one attempted to count the sudden jars and prolonged throes, so rapid was their succession. And even during the intervals between the quakes, the ground and all objects upon it seemed to quiver like the surface of a boiling pot. The quaking was most fearful in Kau. . . . The shocks and quiverings cintinued with different degrees of intensity until Thursday, the second inst. [April] . . . when, at 4 P. M., a shock occurred which was absolutely terrific. All over Kau and Hilo the earth was rent in a thousand places, opening cracks and fissures from an inch to many feet in width, throwing over stone-walls, prostrating trees, breaking down banks and precipices, demolishing nearly all stone churches and dwellings, and filling the people with consternation. This shock lasted about three minutes.

Mr. F. S. Lyman[8] writes as follows of his experiences at Kau during this disturbance:

First the earth swayed to and fro from north to south, then from east to west, then round and round, up and down, and finally in every imaginable direction, for several minutes, everything crashing around and the trees thrashing as if torn by a hurricane, and there was a sound as of a mighty rushing wind. It was impossible to stand; we had to sit on the ground, bracing with hands and feet to keep from being rolled over. . . . The villages on the shore were swept away by the great wave that rushed upon the land immediately after the earthquake.

Some observers estimated that more than 2,000 shocks occurred during this period of disturbance. In spite of the violence of this earthquake on Mauna Loa, it was quite local in extent. No damage was done in the northern half of Hawaii even by the heavy shock of April 2. This shock was felt distinctly on the island of Maui, 110 miles distant, for 90 seconds, shaking furniture, pictures and walls and causing small sea waves. At Oahu, 210 miles from the centrum, the shocks were slight, and though they occurred in the middle of the afternoon, most of the inhabitants of Honolulu were not aware that an earthquake had occurred.

From the human standpoint, the most disastrous of the earthquakes assigned to volcanic causes is that which occurred at Casamicciola on the Island of Ischia, July 28, 1883. When it took place there was a large assemblage of people in the theater, which was of stone and collapsed under the shock, killing most of the audience. Only one house in the whole town was left standing and it is estimated that about 1,900 people lost their lives in the disaster. In

[8] *Am. Jour. Sci.*, II., XLVI., 110, July, 1868.

Naples, however, only twenty-two miles away, the shock was felt by but few people, and the seismographs in the observatory on Mt. Vesuvius did not record it at all, though the instruments at Rome and Florence showed the passage of some extremely light vibrations. The depth of the focus has been calculated at about a half mile and Casamicciola received the vertical shock. The latest eruption of Mte. Epomeo, Ischia's great volcano, occurred in 1302.

Many other instances of volcanic earthquakes might be cited, but perhaps none within historic times have been more severe than those which have been mentioned. All show extremely restricted areas of disturbance, a fact which indicates a comparatively slight depth for the origin of the shocks and a far smaller expenditure of total energy than is developed in connection with the great tectonic quakes. It must not be overlooked, however, that some earthquakes, the origin of which is doubtful, may rightly be assigned to a volcanic origin. Furthermore, the intrusion during past geologic time of countless dikes, sills and laccoliths of igneous rock, the occurrence of which is known from exposures all over the world, must have been accompanied by sudden dislocations, causing earthquakes. Such quakes would be of combined volcanic and tectonic origin. It cannot be asserted positively that they are not occurring at the present epoch.

This brings me now to the consideration of the third and most important class of volcanoes, viz., tectonic quakes, or those which are caused by dislocations in the earth's rock crust due to the action of mountain-building forces. Mountain regions of high geological antiquity, like the Appalachian protaxis and the Scandinavian Peninsula, have had time to adjust themselves to the crustal strains due to their elevation and hence are rarely the scene of great earthquake shocks. In the younger mountain systems, however, such as the Apennines, the Japanese archipelago, Central America and those of California, where young strata abut unconformably against old, the adjustment to strains is still going forward, the cumulative effect being followed by sudden and irregular release of pressure, producing the vibrations which we know as earthquakes. Some of these tectonic quakes have sensibly affected enormous areas. That of Lisbon, 1755, was felt from northern Africa on the south to

Scandinavia on the north and to the east coast of North America on the west, an area estimated by Baron von Humboldt at four times that of the whole of Europe. The Andean earthquake of 1868 shook severely a strip of country 2,000 miles long. The modern seismographs have given pronounced records of earthquakes whose origin was certainly not less than 8,000 miles distant—truly world-shaking events.

The depth of the origin of the shocks below the surface of the earth probably never exceeds thirty geographical miles and usually is not more than from five to fifteen miles. The geological structure of the region through which the earth waves are propagated affects the rate of advance of the same earthquake in different directions and produces many changes in the direction of movement and great differences in the destruction wrought upon buildings. Heavy earthquake shocks are transmitted through the earth at a greater velocity than light ones and the same shock shows different rates in different materials.

In the case of distant quakes three disturbances are recorded instrumentally. The first set of waves to arrive comes on a direct course through the earth's mass; the second set comes along the shortest route on the surface, while the third set arrives by the opposite and longest surface route. The last are comparatively feeble, and they may arrive three and one half hours behind the second set. The first set of waves, those coming through the earth, are propagated with the greatest velocity, which is practically uniform and is about ten kilometers (6¼ miles) per second. These direct waves have been shown by Marvin to be longitudinal in character, and this character combined with their velocity is supposed to cause them to give out the musical sounds which are the premonitory rumblings of an earthquake. The second set are the surface waves due to the "principal portion" of the earthquake, and the increased use of delicate instruments of measurement has led to the acceptance of 3.3 km. per second as their normal rate of propagation. The determination of these various velocities leads to the conclusion that the crust of the earth is practically uniform in constitution to a depth of at least thirty miles.

The duration of an earthquake and the number of shocks in it

vary indefinitely. The Charleston, San Francisco, Kingston and many other quakes lasted only from thirty to forty seconds. Milne states that the average duration of 250 earthquakes of moderate intensity recorded by instruments in Tokyo between 1885 and 1891 was 118 seconds. The first shocks are almost always succeeded by after shocks which may continue for weeks, months or even years.

It has not been possible yet to determine the periodicity of shocks or to predict with any degree of accuracy the time of the occurrence of an earthquake. Some earthquake regions are subject to frequent shocks, while others experience them only at long intervals. The frequency of earthquakes, considering those of all amplitudes, is not generally realized. The globe, indeed, may be said hardly ever to be free from seismic disturbances of some kind somewhere, since the average of all recorded shocks, according to de Montessus de Ballore, is more than fifteen per day, and there are between fifty and sixty heavy shocks per year. The bare enumeration by this author of those occurring in 1903 alone fills a book of six hundred tabulated pages, and he has compiled the data and plotted the position of 159,781 earthquakes that have been recorded up to the end of 1903.

At the same time that important quakes are the result of tectonic movements in the earth's crust, they may themselves be the causes of more or less important changes in the surface of the earth. Sharp waves passing through mountain regions have been known to produce land slides, shatter rocks, displace larger or smaller segments of cliffs, open fissures in the soil or cause subsidence in alluvial regions. Springs, brooks, rivers and lakes have been formed, altered or obliterated as a result of earthquake action. Great earthquakes have usually produced important sea waves causing much destruction along the coast and, sometimes, permanent changes due to erosion and transportation of material.

Several scales for the purpose of indicating the severity of an earthquake shock have been proposed. The one most commonly employed is known as the Rossi-Forel scale, which distinguishes ten degrees of intensity according to the effects produced upon human observers and structures. Another widely used scale is that which has been devised by Professor G. Mercalli. This likewise

consists of ten degrees of intensity and depends upon human observers and the effects upon buildings for the classification of a shock.

On account of the vagueness of these series, the influence of the personal equation of the observer in placing shocks in accordance with them and the over-importance attached by them to effects upon human property, other scales have been proposed, the best of which are based upon instrumental records. Difficulties in using the latter, however, arise through the small number of instruments actually at work, and the Rossi-Forel and Mercalli scales are still found very useful, particularly in the collection of data.

I shall close what I have to say regarding the subject of the afternoon by brief descriptions with illustrations of the earthquakes that occurred at Charleston, S. C., in 1886, at San Francisco in 1906, at Kingston, Jamaica, in 1907, and at Messina in 1908.

The Charleston Earthquake.

The most important earthquake occurring in the eastern part of North America during the historic period was that which devastated Charleston, South Carolina, in 1886. This was investigated under the auspices of the United States Geological Survey by Major Clarence E. Dutton and his assistants, their report being published in the Ninth Annual Report of the survey.

About eight o'clock in the morning of August 27, 1886, the villagers of Summerville, 22 miles northwest of Charleston, S. C., were startled by the noise and shock of what was at first thought to be a heavy blast or a boiler explosion. The sound seemed very near, but no cause for it was learned that day. Around five o'clock the next morning the noise and shock came again and more heavily, and the idea that an earthquake had occurred became general and was strengthened by light tremors that were felt that day and the next. The affair seemed then to be over, for nothing unusual was heard or felt on the thirtieth and during daylight of the thirty-first. The noises or shocks were felt by very few people in the city of Charleston, but they were the premonitions of the great earthquake

that began at 9:15 P. M. of the thirty-first. In the words of Dr. G. E. Manigault, a resident of Charleston, as quoted by Dutton:[9]

> Although the shocks at Summerville excited uneasiness in Charleston, no one was prepared for what followed. . . . As the hour of 9:50 was reached there was suddenly heard a rushing, roaring sound compared by some to a train of cars at no great distance, by others again to an escape of steam from a boiler. It was followed immediately by a thumping and beating of the earth underneath the houses, which rocked and swayed to and fro. Furniture was violently moved and dashed to the floor, pictures were swung from the walls and in some cases completely turned with their backs to the front, and every movable thing was thrown into extraordinary convulsions. The greatest intensity of the shock is considered to have been during the first half, and it was probably then, during the period of the greatest sway, that so many chimneys were broken off at the junction with the roof. The number was afterwards counted and found to be almost 14,000.

Apparently there were two maxima, the first of ten seconds duration, the second of six, with an interval of comparative quiet of 22 to 24 seconds. The whole period to be assigned to this destructive double shock is about 68 seconds.

Another observer states that four severe shocks occurred before midnight and that three others followed at about 2, 4 and 8:30 o'clock A. M.[10] Afterquakes occurred for months. Twenty-seven persons were killed outright and at least 56 more died from injuries received and exposure suffered. The money value of the property destroyed was estimated for Charleston alone at between $5,000,000 and $6,000,000. Not a building wholly escaped injury. Damage to buildings was greater on the low made ground than on the natural higher parts of the city.

The occurrence of visible surface waves was so definitely asserted by so many observers and with such detail of description that the fact of their formation cannot be discredited. The passing of such waves has often been included in the description of earthquakes, but their actual existence had been doubted, on account of the difficulty of explaining their origin. The amplitude of the surface waves in some parts of Charleston is estimated by Dutton at nearly or quite a foot and the average amplitude for the city at three or four inches.

[9] Ninth Annl. Rept. U. S. Geol. Survey, p. 231. Washington, 1889.
[10] *Op. cit.*, p. 217.

Besides throwing down walls and chimneys and moving houses bodily on their foundations, the earthquake caused wooden posts and brick piers to sink vertically into the earth; compressed railroad tracks into more or less complicated curves or stretched them apart; opened innumerable fissures in the ground, and formed hundreds or craterlets at many places out of which gushed water, sand and mud in copious streams.

The earthquake waves traversing Charleston were localized as coming from the northwest and from the west. The principal epicentrum was determined as being about sixteen miles northwest of the city and one mile from the little railway station at Woodstock, and a secondary epicentrum about fourteen miles due west of town. The focus of disturbance was a line or plane estimated as being twelve miles below the surface "with a probable error of less than two miles." The velocity of the wave motion throughout the eastern half of the United States was calculated as averaging 190 miles per minute. The intensity reached No. 2 of the Rossi-Forel scale as far away as New Orleans, Clinton, Mo., La Crosse, Wis., Saginaw, Mich., Burlington, Vt., and Boston—an extreme radius of about 1,000 miles. The Charleston earthquake is classed as a tectonic quake, though no evidence of faulting was apparent on the surface.

(Lantern slides were shown depicting the destruction of buildings in Charleston and vicinity and the formation of fissures and craterlets.)

The San Francisco Earthquake.

California has always been known as a seismic region. Professor E. S. Holden has catalogued 514 shocks, 254 of which affected the region of San Francisco alone, within the period between 1850 and 1886. During the nineteenth century there were ten severe quakes; that of 1868, known as the Mare Island quake, having such a disastrous effect upon the city of San Francisco that serious doubts were entertained of the advisability of rebuilding on the same site, but these fears were soon forgotten and the city rapidly rose again. It was rebuilt, however, without much reference to the lessons that might have been learned from the experience.

In the Sierra Nevada, forming the eastern half of the state, earthquakes are likewise frequent. In 1872 occurred the great Owens Valley quake, which was one of the most severe on record and was the result of movements producing a series of faults along a line more than 100 miles long with a throw of from ten to twenty feet. This mountain system is formed of Precambrian granites, gneisses and schists, upon which have been laid down upon the west an unconformable series of late Paleozoic and Mesozoic strata. The coast ranges, in which the earthquakes occur with far greater frequency, are composed of a granitic core against which rest extensive Mesozoic and Cenozoic strata upon which are thick marine Pleistocene and recent beds. The latter are full of the fossil shells of still living species of mollusks and show that elevation is still going forward in California.

The San Francisco Peninsula is traversed by at least five known lines or zones along which movement, or faulting, has occurred again and again. The principal of these zones is the San Andreas, which takes its name from an important lake through which it runs. It is likewise known as the Stevens Creek fault, as the Portolá-Tomales fault or more simply as "the rift." This zone continues northwest in a slightly curved line to Point Arena and southeast to the mountains west of Hollister. This is the continuous extent of the fault, some 190 miles, but it probably extends under the ocean beyond Cape Mendocino to the north and into the mountains southeast of the line recently disastrously affected.[11] According to H. W. Fairbanks[12] the recognized rift extends from Shelter Cove, Humboldt County, as far southeastward as the Colorado desert and is 700 miles long. Dr. Fairbanks states further that the great Tejon earthquake of 1857 was caused by movement in the same fault zone.

The recurrence of horizontal and vertical movement along the northern 200 miles of this fault line caused the earthquake which at

[11] "The California Earthquake of 1906," by David Starr Jordan and others. G. K. Gilbert, map, p. 317. San Francisco, 1907.

[12] "The California Earthquake of 1906," pp. 321–337. See also "Report of the California State Earthquake Investigation Commission," by A. C. Lawson, chairman, p. 48. Washington, 1908.

5:12 o'clock A. M., western time, April 18, 1906, wrought ruin or serious damage over a belt 50 miles wide and 300 miles long. The approximate position of the epifocal point of the disturbance is given by F. Omori as being in latitude 38° 15′ N. and longitude 123° W., near Tomales Bay.[13] The horizontal shearing movement varied from nine to twenty feet toward the N.N.W. or the S.S.E.; the vertical movement did not exceed two feet at any locality and usually was absent, upthrow where present being on the west side of the rift. Among the effects along the line of the fault were rifting and bulging of the soil, offsetting of fences, roads and walks, splitting and overturning of trees, landslides in the mountains, wrecking of railway tunnels, spreading and telescoping of lines of waterpipe. This is the most disastrous earthquake that has visited the United States, though the chief destruction wrought was due to the fire that followed in the train of the quake rather than to the shock itself. About four hundred people are known to have lost their lives in the catastrophe, and at least $350,000,000 worth of buildings and other property were ruined by the shock or consumed by the flames. An exact statement of the pecuniary loss caused by the shock cannot be made, but the insurance companies finally agreed upon a settlement to the effect that one-fourth of the damage was due to the earthquake and three-fourths to the fire, and this estimate may be accepted as the best that can be made. More than four square miles of the city of 400,000 inhabitants was devastated.

The main part of San Francisco lies about eight miles northeast of the fault line, and the propagation of the waves through the city was in a direction N. 76° E., nearly normal to the fault line. In general the advance of the wave motion on each side of the rift was away from it. Omori concludes that both sides of the fault line were displaced toward the N.N.W., the west side more than the east, the amount of apparent slip being merely differential. In San Francisco the chief damage was wrought upon structures built upon alluvial or made ground. High steel-frame structures which were not stiffly braced acted like inverted pendulums, causing ruin to their walls. This was illustrated in the case of the City Hall in San Francisco and the library buildings at Stanford University

[13] "The California Earthquake of 1906," p. 289.

and the City Hall at Santa Rosa. The main source of the earthquake is thought to have been situated at a considerable depth below the surface (Omori).

(Lantern slides were shown to illustrate the destruction of buildings in San Francisco, Santa Rosa and Leland Stanford Jr. University, and the geologic and topographic changes wrought in the surface of the ground along the line of fracture.)

The Kingston Earthquake.

The Blue Mountains, rising 7,400 feet above the level of a sea 18,000 feet deep, form the back-bone of the island of Jamaica. They trend northwest-southeast and, according to Robert T. Hill,[14] from the earliest axis of folding now apparent. Upon this have been super-imposed later east-west flexures corresponding with the crustal movements that early in the Mesozoic era determined the chief characteristics of the Greater Antilles. Charles W. Brown,[15] reports observing "transverse faults in the Blue Mountain region which undoubtedly indicate lines along which fractures may occur." Professor Hill assumes an east-west axis of folding with an anticline producing the trend of the Greater Antilles and leaving a parallel syncline coinciding with the Bartlett Deep just north of Jamaica.

Such strong relief coupled with folding indicates a high state of tension in the earth's crust. Resistance to stress is diminished on steep slopes, especially when the application of pressure to the ends of an axis is not made in the same plane, giving rise to torsional strains. Fracturing results, tending to follow old fault planes, and these fault planes were originally determined by zones of weakness in the rocks. Fracturing, as we have seen, produces earthquakes. Montessus de Ballore acquiesces in the folding postulated by Hill and embraces the Greater Antilles, including Jamaica, within the great Alpine geosynclinal. The region experiences frequent shocks and one of the most dreadful disasters of modern times occurred within it in the year 1692, when, as a result of an earthquake, the greater part of Port Royal, the capital of Jamaica, sank into the

[14] *Bull. Mus. Comp. Zoöl.*, Vol. XXXIV., p. 164.
[15] *Popular Science Monthly*, Vol. LXX., p. 385, May, 1907.

sea. The city was built upon a narrow sand spit formed of the detritus brought down by rivers from the mountains of the interior or cast up by the sea. It is estimated that 2,000 people lost their lives in this disaster, when a tract of land about a thousand acres in extent sank so as to lie thirty or forty feet under water.

After the destruction of Port Royal the city of Kingston was established on the gradually rising Liguanea plain across the harbor from the old capital, and it flourished for 215 years, becoming a compact city of 60,000 inhabitants. Its business portion extended along the water front and was only twelve blocks long and two wide. The city was built, however, upon unconsolidated gravels and sands —alluvial and coast deposits that gave a foundation but little more secure than the sand spit gave to old Port Royal. Hence when the earthquake of January 14, 1907, occurred, 85 per cent. of the buildings in the city was injured or destroyed, and fire completed the ruin over ten or fifteen blocks of the business and warehouse section.

The shock probably began at 3:33 P. M., though an exact statement cannot be made through lack of accurate standard time in the island. This defect as to time has made it impracticable to plot any coseismal lines. The first series of vibrations, the great shock, lasted 35 seconds, more or less, but the duration varied with the position of the observer. The longest period was reported from the north shore and as being 90 seconds. After the preliminary tremors, which were heard before they were felt, the shock was double, the first maximum being reached in about ten seconds, followed by a second and less acute climax before the vibrations ceased. The interval between the preliminary tremors and the main shock was almost insensible. After shocks occurred for several months. Through the city of Kingston and its immediate vicinity the earthwave shown by the first climax passed from west to east, but three miles north of town the direction of motion was distinctly from the south, while in the Hope River valley five miles east of the city, the advance was from the northwest. The earthwave recorded by the second maximum of shock was more undulatory in character than the first and seems to have originated more to the south of the city. This direction of motion combined with the first produced a twisting counter-clock-wise movement of slender upright structures

like statues, columns and chimneys and had a noticeable effect on buildings.

According to Professor Brown:

> The dip of the angling cracks at Kingston points to a locus of disturbance much to the west of that city, while the lines of isoseismals indicate the intensity area to be in the eastern half of Kingston. . . . The only conclusion then is that the eastern end of the Liguanea plain was the nearest area to the real epicenter that by nature of material would give the greatest amplitude to the destructive epifocal waves. Further, the angle of emergence at Kingston coördinated with the proximity of a probable epicenter together with the limited area of disturbance indicates a shallow origin of about three miles.

As is demanded by theory and observed in fact the vibrations increase in violence on passing from an elastic to an inelastic medium—the destruction wrought in Messina, San Francisco and other places has been worse in the sections built upon alluvial or other loose soil than in those built upon rock, and Kingston was entirely upon such loose material. The experiences of these and other regions show that the destructiveness of an earthquake is not necessarily greatest in the epifocal area. If the locus of disturbance is in or under an elastic rock-mass and the shock is propagated into a region of inelastic loose material, the destruction in the latter may exceed that in the real epicenter. The fault which was the locus of the San Francisco quake is some miles from the city.

The shock of the Kingston earthquake was not sensible on the island of Haiti to the east or on Grand Cayman to the west, but Santiago de Cuba, 120 miles to the north, felt it slightly. This indicates an ellipse as being the generalized form of curve for the isoseismals, with the longer axis extending approximately north and south. At Annotta and Buff Bays on the north shore of Jamaica, opposite Kingston, the destruction wrought was almost as severe as at the capital city. The inference is that renewed faulting along north-south fault lines caused the earthquake.

The building construction of Kingston was as bad as the foundation upon which the city rested. Brick structures predominated, but for the most part it was evident that the brick had been laid dry in poor mortar. Such buildings collapsed under the shock. Those that were properly put together withstood the quake better. Wooden houses with good braces and well fastened together were not thrown

down. Massive walls showed cracks from half an inch to two inches wide. The double amplitude of the wave motion of the earth is estimated at not more than one inch. Such an amplitude is small when compared with the four-inch amplitude calculated by Omori[16] for the earthwave of the San Francisco (1906) quake, the 6 to 12-inch amplitude estimated by F. A. Perret[17] for the earthwave at Messina in last December's quake, or the one foot maximum amplitude given by C. E. Dutton[18] for the Charleston earthquake wave. These largest estimates were derived from effects in soft ground and are probably excessive.

From a geological standpoint the movements causing the Kingston earthquake were less important than the changes in the earth's surface that were produced by it. Surface evidence of the former has not yet been discovered, but the latter are quite apparent. Beginning in the city water front, a belt of fissuring and subsidence skirted the eastern half of the harbor and returned along the inner (northern) base of the Palisadoes. Opposite the city the zone of disturbance forked, one branch maintaining the original direction and passing through Port Royal, while the other curved northwestward touching Ft. Augusta and dying out in the River Cobre valley, eight to ten miles northwest of town.

From soundings taken for Professor Brown, it was learned that "in several places along the edge of the harbor the bottom had sunk from old soundings of a fathom and a half to over six fathoms, and that on the harbor side of the base of the Palisadoes a series of step faults reached a maximum depression at the shore to the north of four fathoms." Port Royal sank from 8 to 25 feet. The zone of disturbance was from 100 to 300 yards wide, containing where exposed many fissures and craterlets out of which water, sand and mud gushed to heights of three or four feet. The fissuring was caused by the compression and expansion of the earth due to the passage of the earthquake wave, but the cause of the subsidence is not clear, for the harbor as a whole did not sink—only an encircling belt. Perhaps solution of the soft limestone where

[16] "The California Earthquake of 1906," p. 307, 1907.
[17] *Am. Jour. Sci.*, IV., XXVII., 327, April, 1909.
[18] Ninth Annual Rept. U. S. G. S., p. 269. Washington, 1889.

the ground waters enter the harbor left caverns into which the overlying material was shaken by the quake (Brown). No sea wave of importance accompanied or followed the shock.

(A series of lantern slides was used to show the destruction caused in the city, the sinking of Port Royal point and the faulting, fissuring and formation of craterlets along the Palisadoes.)

The Messina-Reggio Earthquake.

Time after time during the historic period Italy has suffered from the effects of serious earthquakes, but never before so severely as from that which occurred in Calabria and Sicily on December 28, 1908, when 200,000 human beings are supposed to have lost their lives. The cities of Messina in Sicily and Reggio in Calabria were completely wrecked, and many other villages and towns were laid in ruins or damaged throughout an irregularly elliptical district 85 miles long by 50 miles wide, extending from Pizzo, Calabria, on the northeast to Riposto, Sicily, at the sea base of Mt. Etna, on the southwest. The epifocal area was the Strait of Messina, with the epicentrum at or near the northern end of the Strait. More precisely, the longer axis of the ellipse of greatest destruction (from Ali to Palmi, about 35 miles), as shown by isoseismals, lies in the strait and runs N.N.E.–S.S.W.

Calabria and northeastern Sicily form a district of extreme seismicity that has been visited by several disastrous earthquakes, among which those of 1783, 1785 and 1905 stand out with prominence on account of their destructiveness to human life and property. Volcanic quakes have been associated with eruptions of Mt. Etna, but they have been strictly local in effect, and their influence has not been seriously felt across the Strait. All the severe shocks have originated in Calabria or under the Strait of Messina are of tectonic character, the geological structure being particularly favorable to the production of such quakes. Forming the backbone of Calabria and extending beyond Messina in Sicily there is an elongated area of Archean gneisses and mica schists. Along this axis there occur nearly horizontal beds of Miocene age up to an altitude of 3,300 feet above the sea, while along the Strait of

Messina there runs a fault with thousands of feet of throw, the uplift being upon the Calabrian side of the Strait. Movement appears to be still going on along this and other fault zones, resulting in repeated earthquakes. Furthermore, the slopes into the submarine depths on both sides of the "toe" of Italy are very steep and therefore unstable.

Toward the end of 1908 the seismic activity of the region was evidently on the increase, and noteworthy shocks were felt November 5 and December 10, while F. A. Perret[19] reports that at 5:20 A. M., December 27, just twenty-four hours before the occurrence of the great shock, the seismograph at the Messina observatory registered an important earth movement. The observatory was wrecked by the great earthquake, but the instruments had been installed in its cellar and Dr. E. Oddone[20] of the seismographic service found them intact and the records intelligible, when he reached the place January 1. These records showed that the quake began at 5:21:15 o'clock A. M., December 28, with a gentle movement the force of which increased during ten seconds and then diminished during ten seconds. After two minutes of calm came the great shock, lasting 30 to 35 seconds, which was recorded by seismographs all over the world. This was followed by comparatively light shocks at 5:45, 5:53 and 9:05 o'clock A. M. of the same day, and by noteworthy quakes at 2:51 and 7:30 o'clock P. M. of the following day. For several days and even weeks minor shocks continued to occur. Some of these "after-shocks" were strong enough to add to the damage caused by the principal quake. According to Mr. Perret[21] the intensity within the megaseismic area was between the ninth and tenth degree of the Mercalli scale decreasing rapidly with increasing distance from the epicentrum, and the centrum was not deeply located, being possibly fifteen kilometers ($9\frac{3}{8}$ miles) beneath the surface.

Messina was a beautiful city stretching for miles along the shore of a magnificent harbor. Lying in an advantageous position on the short cut from the Eastern Mediterranean to the Tyrrhene

[19] *Am. Jour. Sci.*, IV., XXVII., p. 321, April, 1909.
[20] *La Nature*, XXXVII., 103, January 16, 1909.
[21] *Loc. cit.*, p. 321.

Sea, the city has enjoyed prosperity for centuries, in spite of frequent visitation from earthquakes. The city was almost completely destroyed by a shock in February, 1783, but the people seem to have learned nothing from their experience with an unstable land. The Messina of yesterday—the city does not exist to-day—was constructed of stone and rubble and old cement. The buildings lined narrow streets and were three, four and even five stories high with massive walls. Hence when the shock came and raised and then dropped the ground for half a minute, the houses, stores, hotels, churches and government buildings were shaken into unrecognizable heaps of débris, filling the sites of the structures and obliterating the streets. The sea-wall in front of the city was partly destroyed, and the promenade along the harbor sank in places below the water.

Reggio di Calabria likewise has suffered frequently from earthquakes, but until within the past few years the inhabitants had not profited by experience to put up earthquake-proof buildings, and all the old houses in the city were demolished by this latest quake. New houses not more than ten meters (33 feet) high are said to have resisted the shocks perfectly. Throughout the Calabrian earthquake district the buildings erected since the disaster of 1905, according to the specifications of the Milan Committee, are reported to be intact in spite of the severe shaking thus received, but all these are low structures.

Photographs show that there was some fissuring of the ground at Messina, and it is reported that "vast chasms" were opened at both Messina and Reggio, but the latter statement is probably incorrect. Professor G. B. Rizzo is quoted as stating[22] that the sea bottom rose in some places, for he saw several boats out of water at the places where they had been anchored some distance from the original shore. The extensive breaking of telegraphic cables indicates submarine disturbance, but the fact of any considerable change in the configuration of the sea bottom remains to be proven and can only be established by careful soundings. No changes in the coast line have occurred, as far as can be detected without an

[22] *Nature*, Vol. LXXIX., p. 289, January 7, 1909.

instrumental survey. It is stated positively that the ground sank in several places in Messina, Reggio and elsewhere, particularly along the harbor front in Messina and along the sea front and in the center of Reggio; but all the low-lying parts of the two cities were built upon unconsolidated alluvial and shore material, permitting, as in the earthquakes of San Francisco and Kingston, severe and destructive oscillations and displacements.

As is usual with shocks occurring along or near the seacoast, the earthquake was accompanied by a "tidal wave," the sea retreating for a considerable distance and then returning into the strait with growing force. The wave was not at all violent in the deep water of the strait and was of importance only as it came into the shallower water near shore, where it was eight or ten feet high. Its crest swept across the marina, or esplanade, bordering the harbor at Messina two or three minutes after the great earthquake shock occurred, and some comparatively slight damage is assigned to the water. The wave was somewhat higher at Reggio than at Messina and attained its maximum on the coast south of Taormina (Perret). In Reggio the buildings on the low land along the coast were flooded. The wave injured a few boats at Syracuse near the southeastern corner of Sicily; but it was scarcely perceptible at the Island of Malta, about 165 miles south by east of Messina, where it arrived at 7:15 o'clock A. M. The sea gauge at Ischia, about 190 miles north-northwest of Messina registered maximum oscillations of 22 centimeters (8.6 inches) at 2:30 o'clock P. M. and at 8 o'clock P. M. If the former was due to the quake that destroyed Messina and Reggio at 5:25 o'clock that morning the rate of advance northward was much less than it was southward.

(A series of slides was shown illustrating the effects of the earthquake in Messina, Reggio di Calabria and Scylla.)

THE EVOLUTION AND THE OUTLOOK OF SEISMIC GEOLOGY.

(PLATES XV AND XVI.)

BY WILLIAM HERBERT HOBBS.

(*Read April 24, 1909.*)

CONTENTS.

PART I: EVOLUTION OF SEISMIC GEOLOGY.
- Introduction.
- The Natural Development of Seismology Prevented by False Theory.
- The Process of Averaging in Mapping Isoseismals and Coseismals.
- The Evolution of the Fault-block Theory of Earthquakes.
- The Relation of Earthquakes to Volcanoes.
- The Mesh-like Distribution of Volcanic Vents.
- Volcanic Extrusions in Relation to Block Adjustments.
- A Possible Explanation of Volcanic Earthquakes.
- The Conditions of Strain during the Growth of Block Mountains.

PART II: THE OUTLOOK OF SEISMIC GEOLOGY.
- The Ultimate Cause of Earthquakes.
- Earthquake Forecasts.
- Periodicity of Earthquake Cycles.
- Possibilities of Future Prognostication.
- Need of Expeditionary Corps.
- A Service of Correlated Earthquake Stations.
- Preparation of Maps of Fracture Systems.
- Maps of Visible Faults and of Block Movements for Special Earthquakes.
- Rate of Mountain or Shore Elevation by Quantitative Methods.
- Investigation of Earthquake Waterwaves.
- Conclusion.

PART I: THE EVOLUTION OF SEISMIC GEOLOGY.

Introduction.—Speaking generally, the present condition of a science is so largely the consequence of an evolution by slow stages, that if the past be reviewed the present stands revealed. Zoölogy, which began with the encyclopædists as a descriptive science, passed into the comparative stage with the advent of Cuvier, and entered

upon its fruitful genetic period when the modern view-point was given it by Darwin. Looking back upon this evolution, we note that the order is in every way a natural one. The facts of observation should first of all be assembled; they must next be compared with a view to establishing correspondences, and, finally, the explanation of the correspondences must be sought in genetic relationships.

Of geology it may be said, that the natural order of its evolution was exactly reversed; for the genesis of the earth and the full order of events in its history had supposedly been given to man through divine revelation. The growth of the science began, therefore, only after a measure of emancipation from the tyrrany of religious dogma had been achieved.

The Natural Development of Seismology Prevented by False Theory.—It may well be doubted if there is another branch of science which has been so long held in fetters by false theory as the branch of geology which treats of earthquakes. Had fate been more kind, it might have been the earliest to develop; for the seats of ancient culture were in earthquake countries, and it will hardly be claimed that the phenomena of earthquakings are not such as to attract the attention. Theories of cause do, indeed, date back before the beginning of the Christian era, the dominating one being that of Aristotle which connected the quakings with explosive sources of energy, conceiving that gases confined in subterranean cavities brought on quakings in their struggles to escape. For the times, this theory seemed to be well supported by facts, since earthquakes were generally manifested at the time of great volcanic eruptions, and volcanoes and earthquakes were common to the same countries. The Aristotelian theory of earthquakes acquired prestige from the adhesion to it of Strabo and Pliny among the ancient philosophers, and at the opening of the nineteenth century, through its adoption by von Humboldt and von Buch, who then dominated the field of geological thought.

The middle of the nineteenth century is a turning point in the history of nearly all sciences toward a greater exactness of observation. Academic discussions in large measure gave place to careful and painstaking observation or to laboratory experimentation. Yet almost at the moment when Darwin and Huxley were

opening a new world to students of biology, the way to progress in seismology was effectually closed through the commanding authority of a pseudo-scientific work of great compass, written by the English physicist, Mallet. Darwin's great theory was an induction reached on the basis of extended observations and of meditations with an open mind; Mallet, on the other hand, approached his work firmly intrenched in a preconceived notion which the facts were assiduously, though perhaps unconsciously, twisted to confirm.

Assuming that Mallet's method had been a sound one, his elaborate observations conclusively proved the fallacy of his theory; for instead of pointing to a definite centrum, his results ranged with noteworthy uniformity between depths of 10,000 and 45,000 feet. The history of science furnishes no more striking example of a great monograph wrought out with laborious scientific method and yet absolutely lacking in scientific spirit or judgment, for with a naïve simplicity Mallet drew from his results the conclusion that, "the probable vertical depth of the focal cavity itself does not exceed three geographical miles, or 18,225 feet, at the outside." Nowhere in the two bulky volumes of his report is the possibility of a non-existence of the centrum even raised.

As was true of the famous fallacy of Werner concerning the origin of basalt, it was here the commanding position of the author which gave his theory its authority; and, although the impracticability of his method soon came to be generally recognized, the fundamental idea was destined to survive at least half a century as the standard doctrine of seismology. It was the brilliant system of Huyghens for treating the propagation of wave motion carried over bodily to seismology, which caused it to be so warmly welcomed by physicists and elasticians, to whose care this branch of science was thereafter entrusted. As late as 1899, the depth of the imaginary origin of a particular earthquake was sought by no less than four different methods with results which ranged from 21 kilometers on the one hand to 161 upon the other, these results apparently not shaking the worker's faith in the reality of the earthquake focus.

It becomes ever more clear that men of science discover in the main those facts only which their working hypotheses indicate to be important. For this reason a theory which is largely correct, grows

by elimination of the false and augmentation of the true, whereas a theory essentially false yields nothing, and by discouraging effort bars the way to progress. With the aid of mathematics and by an abundance of exact observation, the more or less occult Aristotelian theory was by Mallet clothed in a modern dress and thus made respectable in the company of the modernized sister sciences. The cause of the earthquake disturbance was by the very nature of the theory hidden so deep beneath the earth's surface as to be removed from direct observation, and was, therefore, a matter suitable only for speculation.

At the opening of the twentieth century, almost fifty years after Mallet had modernized the theory of Aristotle, authors of text-books of geology quite generally disposed of the subject of earthquakes by a treatment of the outlines of the Mallet theory in the compass of a few pages. How generally the investigation of earthquakes was excluded from the field of research in geology is strikingly shown by the activities of the United States Geological Survey, a bureau employing the largest staff of working geologists of any in the world and including in its field subjects as diverse as paleontology and mineral resources. In the years 1868, 1872, 1886 and 1887 earthquakes of the first magnitude wrought damage to property within the national domain, and with one exception no effort was made by the national bureau to investigate these phenomena, and but little by independent geologists. Since the intellectual shock from the California earthquake of 1906, individual geologists have begun to take advantage of this opportunity for study, even though the golden opportunity had already passed.

The Process of Averaging in Mapping Isoseismals and Coseismals.—Aside from its occult and speculative basis, which removes it from the reach of direct observational studies, the centrum theory has yet assumed to adopt the observational method of modern science. The isoseismal and coseismal lines which belong to the Mallet conception of an earthquake centrum must be obtained through averaging the results of observation either of the intensity of the shocks or of the time of their arrival. In how far it has been necessary to "adjust" data in order to make the circular or elliptical curves concentric about the epicenter and represent uniformly de-

creasing values as they recede from it, one who has not compared the individual data will scarcely believe. A local intensity which is too large can be explained either by a soft or a wet basement, by an earthquake "bridge" or by probable error of observation; while one too small may be explained by an earthquake "shadow," by an interference of waves, etc. Many curiously anomalous data not possible of explanation on any of these grounds may be dismissed as "earthquake freaks."

As regards time of arrival of shocks, "too early" or "too late" data have not uncommonly been included among those which seemed *a priori* the most reliable. Especially good examples of such data are furnished by the studies of the Agram earthquake of 1880, the Andalusian earthquake of 1885, the Charleston earthquake of 1886, and the Indian earthquake of 1887. Out of 260 time data collected by Dutton in connection with the Charleston earthquake, 47 were rejected as "too early."

To average the determinations of an unvarying value in order to eliminate the errors of observation and experiment, is indication of a desire to secure accuracy which must be commended as eminently scientific in its nature; but to average the values of a property the distribution of which either in space or in time is likely to be significant, is, on the contrary, one of the most pernicious, as it is one of the most common and unconscious methods. Such a practice is often condoned on the ground that the data may otherwise appear to possess an accuracy beyond what they really have; forgetting, what is far more important, that through the averaging process the data lose their most significant characters. Now that so many sciences are entering upon their quantitative stages it is important that this method be corrected.

A companion fallacy to the supposed necessity for averaging data of different values is that nature in all its moods has avoided angles and straight elements in favor of the curving outline, and that in consequence results are incorrect in proportion as they bring out strong accent, or definiteness of character, or exhibit straightness of contour. In no field, perhaps, has this fault been more often committed than in topographic mapping, where it has been encouraged as tending toward accuracy. A new era is dawning, however,

and the wonderfully improved maps which have been brought out in recent years by the United States Geological Survey and by European surveys have been secured through the elimination of the process of averaging and "rounding off" of angles. Significant character is thus taking the place of a lack of expression in the older maps.

In a similar way the isoseismals and coseismals, which have assumed to represent the distribution in space and in time of the seismic activity of a district, have through averaging of results removed all true expression of seismic distribution. It is likely, however, that this method will yet, at least for a number of years, effectually retard the natural progress of seismology.

The Evolution of the Fault Block Theory of Earthquakes.—It would be incorrect to state that no progress was made in seismic geology during the last half of the nineteenth century, but it would be only the truth to say that such progress as there was, was achieved in spite of and almost in defiance of the orthodox doctrine of seismology. Nine out of ten reports upon special earthquakes made during that period have included only the maps of isoseismal and coseismal lines, to which has been added a computation of the depth of the supposed origin.

It is now proposed to trace the development of the tectonic conception of earthquakes as it has grown into the fault-block theory of the present day. To the Austrian school of geologists and to its leader, Eduard Suess, must be credited the pioneer work upon the geology of earthquakes. The discovery of the localization of heavy shocks along definite lines, or the recurrence of epicenters (surface loci of heavy shocks) along such lines, has been a characteristic of the Austrian method, which dates from a paper published by Suess in 1872. Such lines in the surface, generally approximating either to a right or to a broken line, were in some cases identified with the traces of fault planes and in others were shown with much probability to be the course of such displacements. Here, then, was the first important recognition of the tectonic nature of earthquakes, and, as a consequence, the Austrian school of seismologists has since endeavored to examine earthquakes in the light of the geological structure of the affected region.

It must be regarded as quite remarkable that the recognition of this fundamental fact was reached in Austria, for the opportunities offered by the Austrian field were by no means exceptional. In fact, the great surface faults which have been a feature of great earthquakes in other districts, have there been seldom observed. In New Zealand, for example, accompanying a heavy earthquake in 1856, an area of country comprising 4,600 square miles was suddenly upraised to form a visible escarpment varying from one to nine feet in height. This event was duly described by Lyell, who, in the eleventh edition of his widely read "Principles of Geology" reported this and other similar cases apparently without seeing that they throw any discredit upon the centrum theory.

In 1884[1] Gilbert, in a brief note, explained the earthquakes characteristic of the Great Basin of the western United States as due to the interrupted jolting uplift of the mass of the mountains by vertical thrust. The stresses tending to uplift the range aided by a fissure already in existence, accumulate until they overbalance the starting friction upon the fissure, when through movement the strain is relieved and the potential energy of the system reduced. In a later note published in 1890[2] he showed that during the earthquake of 1872 in the Owen's Valley, California, the ground was moved in strips both vertically and horizontally.

In 1893 Kotô, describing the great Japanese earthquake of 1891, in referring to earlier earthquake rents within the same district said:

> The event of October, 1891, seems to me to have been a renewed movement upon one of these preëxisting fissures—the Neo Valley line of fault, by which the entire region lying to the right of it not only moved actually downwards but was also shifted horizontally towards the north-west for from one to two metres along the plane of dislocation. This vertical movement and horizontal shifting seem to me to have been the sole cause of the late catastrophe.[3]

Without the aid of surface faults, Leonhard and Volz, in 1896, expressed clearly the idea that the Silician earthquake of 1895 was the result of an adjustment among orographic blocks or *Schollen*. Their statement was:

[1] *Amer. Jour. Sci.*, Vol. 27, 1884, pp. 49–53.
[2] Mon. I., U. S. Geol. Sur., pp. 360–362.
[3] *Jour. Coll. Sci.*, Tokyo, Vol. V., 1893, p. 329.

We must, therefore, regard the cause of the earthquake of June 11, 1895, as a movement of the Nimpt complex of orographic blocks, which occurred along the southern and eastern fracture margins.[4]

The great Indian earthquake of 1897 was thoroughly examined from the geological side with results which seem to have afforded indication of the movement of the ground in individual blocks. This, however, was not the theory adopted by R. D. Oldham, who wrote the report upon the earthquake, apparently for no other reason than that it seemed to require an expansion of the affected area. In consequence, the unique hypothesis was offered that the earthquake was due to a movement upon a thrust plane beneath the affected region. The mental attitude of Dr. Oldham is brought out in the following paragraphs from his report in modification of his choice of theory:[5]

Though apparently the most probable this is not the only possible, hypothesis. The surface features of the Assam range, described in the last chapter, are compatible with, in some respects they suggest, the idea that these hills are what the German geologists call Schollengebirge, that is, mountains which have arisen from straight up and down thrusts, instead of from lateral compression, like the Alps and Himalayas. *If this be so, the faults by which the fault scarps are formed would be normal faults,*[6] and so far from there having been any compression, the elevation of these hills would have been accompanied by an extension of the surface. The state of strain, too, which preceded the earthquake would have been one of tension and not compression.

The mechanism of the production of this form of mountain is not properly understood, and a condition of tensile strain in the crust of the earth would be still more difficult to explain, but the fact of the existence of such mountains and structures cannot be gainsaid, so the possibility of the state of tensile strain they imply must be allowed.

If such is the nature of the Assam range, and of the cause of this earthquake, there would be no thrust-plane underlying it, and the focus of the earthquake would have to be regarded as a complex one. That is to say, there would be no general focus, but a number of independent ones, along each fault, and the magnitude of the earthquake experienced would be due to the simultaneous occurrence of a number of earthquakes of various degrees of severity.

Whether we regard the focus as a thrust-plane, or as a network of faults, it practically covered an extensive area.[7] The hypothesis of a thrust-plane

[4] *Zeitsch. f. Erdkunde z. Berlin*, Vol. 31, 1896, pp. 1–21.

[5] R. D. Oldham, "Report on the Great Earthquake of 12th June, 1897," *Mem. Geol. Surv. India*, Vol. 29, 1899, pp. 165–168.

[6] The italics are mine.—W. H. H.

[7] The italics are mine.—W. H. H.

is the simplest to work with, as also the most probable, and it is that which has been adopted in the following pages.

As we shall see, the fundamental difficulty which stood in the way of the acceptance of the *Schollen* idea at the time Oldham was writing, has since been removed by the "distant" studies of earthquakes (see below, p. 285), and the theory of a thrust-plane, which he chose to adopt, has remained without any support in later work.

Additional and important contributions toward the fault-block theory of earthquakes have crowded about the beginning of the twentieth century. In the year 1900, Yamasaki, in describing the great earthquake of northern Honshu, which occurred in 1896, gave as its cause the movement on two visible displacements which opened on opposite sides of the mountain mass.[8]

> Two long lines of fracture were discovered by me to be the cause of the Riku-U. earthquake. . . . They lie on the two sides of the mountain axis of the Central chain, and so this earthquake offers an example of the longitudinal quakes (*Längsbeben*) which but seldom occur.

Thoroddsen, in a report which reached the scientific world first through a German abstract of the year 1901,[9] was able to show that during each of the five heavy shocks of the South Icelandic earthquakes of 1896, a separate block of country had been shaken. These several areas were all included in a low plain walled in by a rampart of mountains, and with a single exception they were contiguous areas which did not overlap.

> Each of the heavy shocks was limited to a circumscribed area which was made evident by a mass of collapsed houses, and from this the earthquake waves were propagated outward in all directions.
>
> . . . . . .
>
> The ground beneath the low plain is probably separated into individual parts and the continued movement on these cross lines [across the main fissures on which the volcanoes of the island are ranged.—W. H. H.], as well as the faults between the individual parts, appear to be the causes of the many earthquakes of this district. If one studies the statistical tables of the ruined houses from each shock [given in Icelandic report.—W. H. H.] it is seen that the individual areas are somewhat sharply delimited; while upon them nearly everything was destroyed, the damage outside was relatively small.

[8] N. Yamasaki, *Pet. Mitt.*, Vol. 46, 1900, pp. 249–255, map.

[9] *Pet. Mitt.*, Vol. 47, 1901, pp. 53–56. The full report had appeared in the Icelandic language two years earlier.

Writing in 1902 Professor John Milne, who has done so much to advance seismology, gave expression to his views upon the cause of the larger and smaller earthquakes:[10]

> The earthquakes to be considered may be divided into two groups—first, those which disturb continental areas and frequently disturb the world as a whole, and secondly, local earthquakes which usually only disturb an area of a few miles radius and seldom extend over an area with a radius of 100 or 200 miles.
>
> These former I shall endeavor to show are the result of sudden accelerations in the process of rock-folding accompanied by faulting and molar displacements of considerable magnitude, whilst the latter are for the most part settlements and adjustments along the lines of primary fractures. The relationship between these two groups of earthquakes is therefore that of parents and children.

Professor Milne's studies of "distant" earthquakes had revealed the fact that the world-shaking earthquakes most frequently occur upon the floor of the ocean.

> When a world-shaking earthquake takes place, and its origin is suboceanic, we occasionally get evidence that this has been accompanied by the bodily displacement of very large masses of material. For example, seawaves may be created which will cause an ocean like the Pacific to pulsate for many hours.

To indicate the grand scale of the mass movements of the crust upon the continental areas, a list of twenty-two larger disturbances was compiled by Milne and the following important conclusions drawn:

> If it can be admitted that world-shaking earthquakes involve molar displacements equal in magnitude to those referred to in the preceding list, . . . then, in the map showing the origins of these macroseismic effects, we see the districts where hypogenic activities are producing geomorphological changes by leaps and bounds.
>
> The sites of these changes are for the most part suboceanic troughs. When they occur, the rule appears to be that a sea becomes deeper, whilst a coast-line relatively to sea level may be raised or lowered. For nearly all the regions of the world where they take place we have geological and not unfrequently historical evidence that the more recent bradyseismic movements have been those of elevation. This elevation, however, only refers to the rising of land above sea-level, while the mass displacements seem to be accompanied by sudden subsidences in troughs parallel to the ridges where rising has been observed. In short, at the time of a large earthquake, two

[10] "Seismological Observations and Earth Physics," *Geogr. Jour.*, Vol. 21, 1903, pp. 2, 9, 11.

phenomena are simultaneously in progress. A suboceanic trough may suddenly subside, whilst its bounding ridge may be suddenly increased in height, and the concertina-like closing of the trough may account for the sea-waves.

Dutton, in 1904,[11] included in his classification tectonic earthquakes, and by supplying data concerning the earthquake of Sonora in 1887 contributed an additional example of uplift *en bloc* of a mountain mass accompanied by a great earthquake. Of this range, the Sierra Teras, he says:

> In other words, the range seemed to have been uplifted several feet between faults on either flank.

Yet the implication in the context is that these observations are hardly decisive, and in a paper read before the National Academy of Sciences in 1906[12] it is made clear that Dutton at this time still adhered strongly to a modified centrum view to which he had contributed in 1889 in his report upon the Charleston earthquake of 1886.

The Dutch geologist, Verbeek, in 1905 published a catalogue of the earthquakes of the island of Ambon in the East Indian Archipelago, together with a full account of the heavy earthquake which caused much damage upon the island on January 6, 1898.[13] His study of the distribution of the damage resulting from the latter quake brought out the fact that the shocks were largely limited to narrow zones on either side of a main fault running in a north and south direction across the island, and to similar zones about three additional faults which cross the first nearly at right angles, the stronger shocks belonging with the first mentioned displacement. Of this north and south zone he says:

> The terrane most disturbed, which one designates "*the pleistoseismic area*" does not here have the form of a circle or of an ellipse, as in the case of so many earthquakes, but that of a long band relatively straight, which shows clearly that we have here to do with a *tectonic quake;* now since we have shown above in the description of the geology that there is at the south of Ambon a fault which is prolonged to the north through Ambon and southward . . . to the southern coast, it is altogether natural to attribute the earthquake to a new dislocation along this cleft or fault of the

[11] "Earthquakes in the Light of the New Seismology," 1904, p. 55.

[12] "Volcanoes and Radio Activity," Englewood, N. J., 1906, p. 5.

[13] R. D. M. Verbeek, "Description Géologique de l'isle d'Ambon," Batavia, 1905, pp. 300–323.

earth's crust. Since the formation of this cleft, which is at least of pre-Cretaceous age, doubtless movements have often occurred which continue even to our time. . . .

In the following year the Count de Montessus de Ballore, who had already become known as a seismologist of reputation by reason of his masterly essay upon the distribution of seismicity over the globe, brought out a comprehensive work entitled "Seismic Geography." In this volume, as a result of the study of no less than 170,000 recorded shocks of earthquake, their distribution within each province was analyzed by new and ingenious methods of combination. In each case the known faults of the district under consideration were discussed, and so far as possible, their relation to the seismic distribution was brought out.[14]

Much the clearest demonstration of the adjustment of portions of the earth's crust as individual blocks, and here by well-demonstrated changes of level, is to be found in a paper by Tarr and Martin upon the results of earthquakes in Alaska in the fall of 1899.[15] Some portions of the coast were found to have been elevated, and other smaller ones to have been depressed. The sea, which here cuts up the district by a number of fiords, permitted the changes of level to be measured by the height of the abandoned shore lines of 1899. In the absence of earlier soundings or of correct maps, the submerged areas were determined with much less precision, though forests now below sea level bear abundant testimony to the local direction of the earth movement. Still older abandoned shore lines, appearing as notches above the raised beach of 1899, proved that the latest elevation is but one stage in the progressive, though interrupted, general uplift of the region. Tarr and Martin's statement of their view is as follows:

Briefly summarizing the inferences which the facts seem to warrant, we conclude that in 1899 there was a renewal of mountain growth, uplifting that part of the mountain front bordering the Yakutat bay inlet to different amounts—7 to 10 feet in the southeast side of the bay, and 40 to 47 feet on the northwest side. This uplift occurred all within a little over two weeks and mainly on a single day (September 10). It was complicated by move-

[14] "Les tremblements de terre; Géographie séismologique," Paris, 1906, pp. 475.

[15] "Recent Changes of Level in the Yakutat Bay Region, Alaska," *Bull. Geol. Soc. Am.*, Vol. 17, 1906, pp. 29–64, pls. 12–23.

ments along secondary fault lines, which produced at least three (and perhaps more) major blocks. . . . The first and largest of these blocks, . . . is apparently titlted upward toward the southwest.

Accompanying this faulting was a minor fracturing apparently due to local adjustments in the tilted blocks. Doubtless this minor fracturing is much more common than our observations indicate, for it was discovered in more than half our expeditions into the interior when we went out of the valleys away from the sea coast.

The evidence accumulated for the tectonic origin of earthquakes and their inseparable connection with the process of faulting in rock strata, has shown that seismology must be considered as a part of tectonic or structural geology—that part, namely, which is concerned with the recent and present-day history of the earth. So soon as this fact receives general recognition, the field of study must be added to that now explored by geologists. For their loss in this quarter elasticians will be more than compensated by the enlarged opportunities which are now offered them for studying earth waves as they are registered at a distance upon the newly devised earthquake instruments.

Recognizing, then, that earthquakes manifest the time of operation of these larger mass movements of the earth's crust which have brought about changes in level as well as changes in horizontal position in connection with faulting, it becomes necessary to place the subject *en rapport* with the latest that has been learned in the wide field of tectonic geology. This treatment of earthquakes as a part of tectonic geology was attempted by the present writer in two monographs published in 1907 in connection with a description of the Calabrian earthquake of 1905,[16] and later, in the same year, in a treatise upon seismic geology.[17]

Having in mind the fact that the traces of fault planes are but rarely exposed to view, and in only a small percentage of cases possible of determination from purely geological studies, the investigation of the Calabrian earthquake was directed toward determining whether, (1) there are lines or narrow zones of special

[16] "On Some Principles of Seismic Geology," with an introduction by Eduard Suess. "The Geotectonic and Geodynamic Aspects of Calabria and Northeastern Sicily," with an introduction by the Count de Montessus de Ballore. *Gerland's Beiträge z. Geophysik*, Vol. 8, 1907, pp. 219–362, pls. 1–12.

[17] "Earthquakes, An Introduction to Seismic Geology," New York, 1907, pp. 1–336.

intensity of shocks, (2) whether these are repeatedly the seat of special danger from successive earthquakes, and (3) whether such lines, if they exist, are expressed in the surface of the country as earth lineaments. The investigation showed that at the time of an earthquake the surface of the country affected is peculiarly sensitized to reveal the courses of hidden faults, which, if thus made apparent, may be designated *seismotectonic lines,* and that strong seismotectonic lines correspond in position to the striking lineaments of the country. In this we find a means of deriving through the study of the topography, the tectonic geology and the seismic history, an imperfect yet none the less a valuable map to display the architecture of each seismic district.

It is a curious illustration of earlier misdirection of effort, that up to the year 1907 no detailed map of the fault system within an area disturbed by destructive earthquake had been attempted. The maps which best display the disposition of adjusted fault blocks were the small-scale charts by Thoroddsen and by Tarr and Martin. In the summer of 1907, at the writer's suggestion, the expert topographer and geologist, Mr. W. D. Johnson, of the U. S. Geological Survey, prepared accurate maps of the surface faults of certain areas disturbed during the Owen's Valley earthquakes of 1872, which maps were published in part during the same year.[18] The sudden changes of displacement on individual faults and the mosaic-like structure of the disturbed region were thus brought out with a clearness and accuracy never before attained.

Seismological science may be said to have suitably celebrated its emancipation from the bondage of the centrum theory, when in 1907 there was published from the pen of the Count de Montessus de Ballore the most comprehensive treatise upon the subject.[18a] This book recognized the adjusted fault block theory as the best available working hypothesis of the science, and with a grasp of the subject which was based upon a lifetime of study, and upon a quite unparalleled knowledge of the literature, earthquakes were so treated as to make the work the one authoritative reference book of the science.

[18] In the author's "Earthquakes," Figs. 23, 45 and 64. More complete maps will appear in a special monograph.

[18a] La Science Séismologique, Paris, 1907, pp. 579.

The common characteristic of all phases of the modern tectonic theory of earthquakes, the evolution of which we have now largely traced, is that the adjustments in position or attitude of sections of the earth's crust are regarded as *the proximate cause and not the effect of the shocks themselves.* So far as molar movements have been recognized by the advocates of the centrum theory, they have been regarded as the *direct consequence of volcanic or explosive shocks emanating from a deeper-seated origin.* Two recent papers of a somewhat speculative nature, prepared by an astronomer, have sought the cause of earthquakes in a leakage from the bottoms of the oceans.[19]

The Relation of Earthquakes to Volcanoes.—As already pointed out, the earliest of the generally accepted theories of earthquakes connected them directly with volcanic action, and this idea has survived in the centrum theory. The tendency of later study has been to indicate that while both betray a certain relationship to each other, this is not often of such a nature as to call for a quick response of the one phenomenon to the other. Regions of volcanoes are subject to earthquakes, yet some of the heaviest earthquakes have affected a region distant from any volcanic vents. Again, most great volcanic outbursts are inaugurated by light earthquakes, but great earthquakes produce as a rule no perceptible immediate effect upon the activity of neighboring volcanoes. Thus, for example, during the late Messina earthquake, which was so heavy about the slopes of Etna, that volcano showed no sympathetic response. Catalogues setting forth the seismic and volcanic activity within any province betray, however, certain periods of years during which both seismic and volcanic activity are at either a maximum or a minimum; though within these periods no close time relation of the one phenomenon to the other is apparent. In short, it would appear

[19] T. J. J. See, A.M., Lt.M., Sc.M. (Missou.), A.M., Ph.D. (Berol.), "The Cause of Earthquakes, Mountain Formation and Kindred Phenomena Connected with the Physics of the Earth," PROC. AM. PHIL. SOC., Vol. 45, 1907, pp. 274–414. "Further Researches on the Physics of the Earth, and especially on the Folding of Mountain Ranges and the Uplift of Plateaus and Continents Produced by Movements of Lava Beneath the Crust Arising from the Secular Leakage of the Ocean Bottoms," *ibid.*, Vol. 47, 1908, pp. 157–275.

PROC. AMER. PHIL. SOC. XLVIII. 192 S, PRINTED SEPTEMBER 7, 1909.

that both earthquakes and volcanic activity are different indications of the operation of a more fundamental geological process—mountain formation, with its concomitant manifestation in changes of level.

Going back in the direction of the ultimate cause of mountain building, we are probably correct in assuming that it is a consequence of the contraction of volume of the planet and the wrinkling of the outer shell, as that shell adjusts itself over the diminished volume of the core beneath. In the past much confusion has arisen from assuming that flexuring has taken place within the outermost shell of the earth, and that the faults discovered are an incident to the folding process *within one and the same set of beds*. Thus we have come to speak of "dip faults" and "strike faults," "longitudinal faults" and "cross faults." Later studies have shown that the processes of folding and of faulting within rocks take place under different conditions of load corresponding to different depths below the surface; and that, therefore, the folding which accompanies the rise of a mountain range is so deeply buried beneath the roots of the range that it can be laid open for study only after a blanketing layer of rock some miles in thickness has been removed. Those mountains which are growing to-day—such, for example, as the Sierra Nevadas of the Pacific border of our own country—are being pushed up in blocks which are outlined by steep faults. The elevation goes on spasmodically, and each successive uplift causes a jolt which is manifested as an earthquake more or less destructive, according as the movement is of large or of small amplitude. Deep below the surface, the rising blocks of the crust rest upon arches of folds which a future generation of geologists may be privileged to study after a layer of the present surface some miles in thickness has been carried away. Those parts of the earth's crust which are not shaken by earthquakes are, in the language of de Montessus, no longer living—they are dead.

Not only are earthquakes the indication of changes in level such as accompany the process of mountain growth, but active volcanoes are now recognized to afford evidence of the same movements. Wherever mountain ranges are now rapidly growing, there active volcanoes are to be found. The full significance of this fact

is only beginning to be appreciated. Fortunately this hypothesis may be fitted to the now quite generally accepted view that the earth is essentially solid throughout, and is maintained in that condition at great depths below the surface by the high pressure from the superincumbent material. Now the arching of strata in the process of folding is competent to lift the load from underlying rocks, so that wherever their temperature is such that fusion would occur at the surface, a reservoir of molten lava is produced and will be brought to the surface from the action of gravity whenever a path is open for it. A reason is thus found for the presence of lava bodies at moderate distances only from the surface in those districts where the process of mountain building is in operation.

The Mesh-like Distribution of Volcanic Vents.—The lineal arrangements of volcanoes and the dependence of this alignment upon the existence of fissures through the crust, seems to have been one of the earliest of geological observations, so soon as the less civilized continents had been scientifically explored. In Europe the systematic arrangement of volcanoes is much less strikingly displayed, and it was there in consequence a later discovery. The credit for having first recognized this important fact of observation is generally given to von Buch, because of his classical study of the Canary Islands. It seems probable, however, that Alexander von Humboldt, his friend and colleague in the field of geological exploration, was the first to make the observation. The latter showed that the volcanoes in the Cordilleran system of South and Central America furnish striking examples of such alignment. Von Buch, in his turn, emphasized this significant relationship, but found certain volcanic districts within which the alignment of vents was not apparent, and so he distinguished *volcanic chains* from *central volcanoes*. Other explorers like Dana and Darwin soon added confirmation of a linear arrangement from the regions which they had individually visited. Dana, a member of the Wilkes Exploring Expedition, brought out the lineal arrangement of the Polynesian Islands and showed that all these were alike rows of partly submerged volcanic peaks.[20] Darwin, during his voyage on the

[20] "Manual of Geology," pp. 37, 282.

"Beagle" made observations[21] which advanced the knowledge of volcanic distribution, as we shall see, very nearly to that of the present day.

As early as 1825, that pioneer and master of vulcanology, Paulett Scrope, discussed the arrangement of volcanoes in the following manner:[22]

> The generality of volcanos have a decided linear arrangement; one vent following the other in the continuation of the same straight or nearly straight line; and when volcanos have been formed on neighbouring points out of this principal line, they are in almost all cases situated upon other rectilinear bands parallel to the first.

Later Scrope expressed his doubt of the existence of v. Buch's class of central volcanoes, for which it had been claimed no alignment could be discovered.[23] In 1844 Darwin proved the existence of neighboring parallel fissues outlined by volcanoes, and was further able to show by his studies of the Galapagos Islands that the arrangement of the vents there brought out the existence of a network of fissures composed of two rectangular series with the principal vents at the intersecting points.[24] The directions of the two series were northwest by north and northeast by east. Virlet d'Aoust had already discovered the same kind of structure in the arrangement of the volcanoes within the Grecian archipelago.[25]

Inasmuch as a mesh-like disposition of volcanic vents within a network is of the first importance in its relation to the mass displacements which occasion earthquakes, it is pertinent to examine the more recent literature of the subject with a view to establishing its truth or falsity. The newer and more accurate methods for preparing maps which have been introduced since the time of Darwin, make such a review at the present time in every way desirable. There are two regions especially which have been recently carefully studied by authorities of the first rank in the field of vulcanology. I refer to Iceland, surveyed at his personal expense throughout a

[21] "Geological Observations on the Volcanic Islands, etc.," 1844, pp. 140–145.

[22] "Considerations on Volcanos," London, 1825, p. 126.

[23] "Volcanos," London, 1862, p. 258.

[24] *L. c.*, edition of 1900, p. 131.

[25] *Bull. Soc. Geol. France*, Vol. 3, 1832–33, pp. 103–110, 201–204.

period of seventeen years by Professor Thoroddsen of Copenhagen, and the islands of the East Indian Archipelago, surveyed for the Dutch Government by the distinguished geologist, Verbeek. Of the Icelandic volcanic region Thoroddsen says:[26]

Of larger eruption fissures and crater chains I have found 87, all of postglacial origin; . . .

. . . The many fissures which are common to several districts can not possibly be entered upon a map of small scale; the terrane is often so divided by clefts that both within the flat country and upon the slopes of mountains it appears to be separated into numerous narrow strips some kilometers in length. . . .

. . . Between the numerous non-volcanic and the volcanic clefts which have poured out important streams, no difference is to be noticed; an ordinary cleft may suddenly become volcanic. . . .

. . . Where larger fissure systems cross, there are often found large

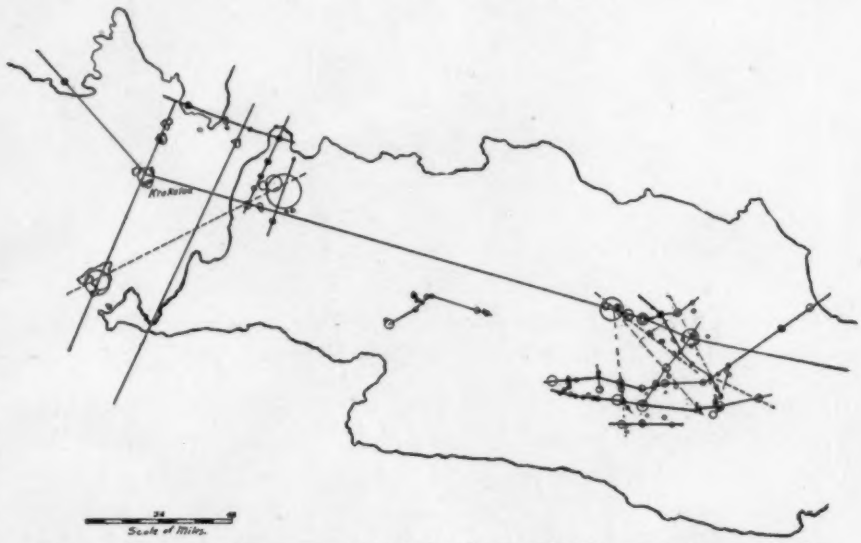


FIG. 1. Map showing arrangement of volcanoes in the western part of the Island of Java. (After Verbeek.)

volcanoes, as for example the largest volcano in Iceland, Askja, with a crater of 55 sq. km. area situated at the intersection of the southland fissure running NE.-SW. and the northland one trending N.-S.

[26] "Die Bruchlinien Islands und ihre Beziehungen zu den Vulkanen," *Pet. Mitt.*, Vol. 51, 1905, pp. 1–5, map pl. 5.

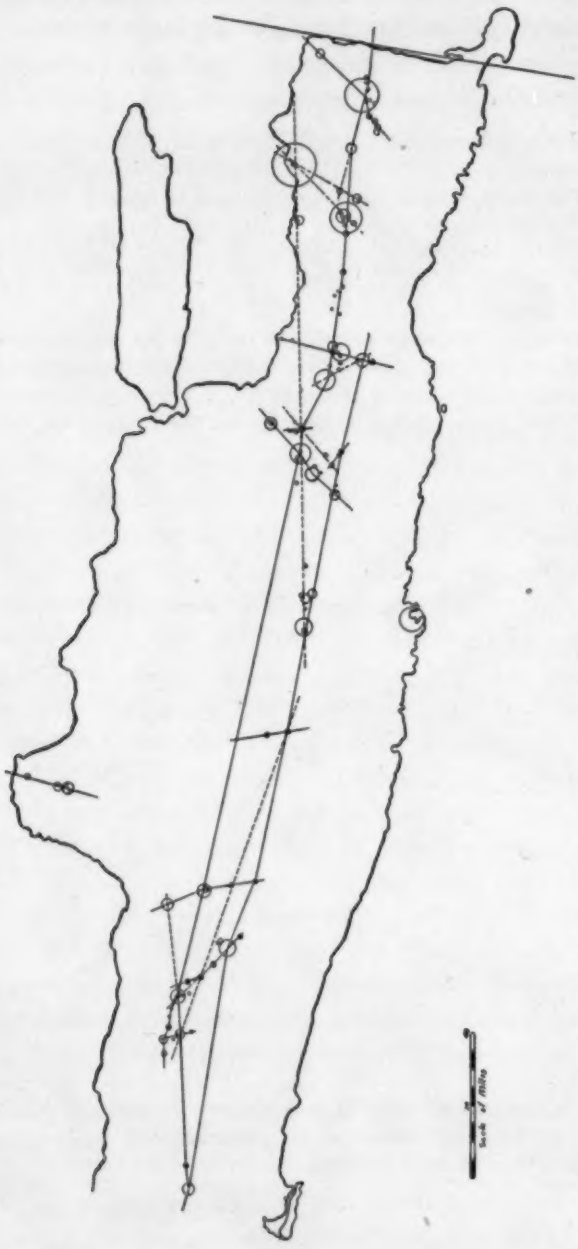


FIG. 2. Eastern portion of the Island of Java showing mesh-like arrangement of the volcanic vents. (After Verbeek.)

With the exception of the report on Krakatoa the five monographs and accompanying grand atlases which have been issued by the Geological Survey of the Dutch East Indies under the direction of Dr. Verbeek, seem to be but little known; yet they contain the results of extended and detailed surveys within one of the world's most interesting volcanic regions.[27] Nowhere have such trustworthy data been compiled which permit of a thorough study of the arrangement of volcanic vents. Clearly aligned upon fissures the map of Java displays the elements in the intersecting volcano network, as may be seen from atlas drawings reproduced in Figs. 1–2.

Though more accurately worked out, it does not appear that these instances of intersection of volcano rows is exceptional. Felix

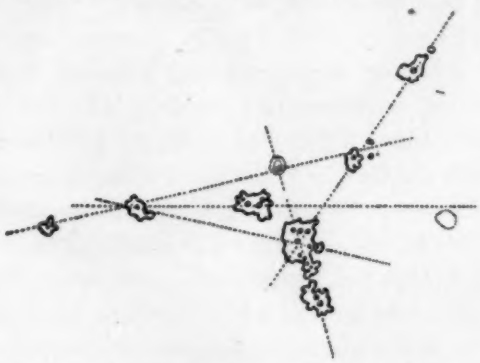


FIG. 3. Map to bring out the arrangement of volcanic islands and submerged volcanic peaks in the Lipari group.

and Lenk[28] have explained the prominence of the mighty volcanoes of Mexico, Popocatepetl, Ajusko and Nevada di Toluca, as due to their location at the intersection of important fissures, though the warrant for this has been questioned by others. The volcanic Lipari Islands of the Mediterranean, which were formerly regarded as

[27] Verbeek, "Sumatra's Westkust" (Dutch language), Batavia, 1883, 674 pp., atlas of 16 maps. Verbeek, "Krakatau," Batavia, 1885, 567 pp., atlas of 25 pls. Verbeek et Fennema, "Description Géologique de Java et Madoura," Amsterdam, 1896, two volumes, 1,183 pp., atlas of 24 maps. Verbeek, "Description Géologique de l'ile d'Ambon," Batavia, 1905, 323 pp., atlas of 10 maps. Verbeek, "Rapport sur les Moluques," Batavia, 1908. 1844 pp., atlas of 20 maps.

[28] *Zeitsch. d. deutsch. geol. Gesell.*, Vol. 44, 1892, pp. 303–326.

built up on radial fissures going out from the ruptured center of a depressed area, reveal a regular plan with the volcanic peaks and craters at the crossing points of intersecting lines, so soon as the submerged cones are brought into the problem (see Fig. 3).[29] The volcanoes of Italy and surrounding waters furnish an example of a much larger network within which the vents are located at intersecting points.[30]

What is true of the arrangement of ordinary volcanic cones within individual provinces, is repeated in the case of the monticules or parasitic cones which are built up upon the flanks of larger composite volcanoes, such, for example, as Etna.[31] To some extent a similar arrangement may be inferred on a far grander scale than any that has been mentioned, as in the longer trains of the volcanic islands. As long since pointed out by Neumayr, the volcanic island, St. Helena, is located at the crossing point of two long lines of widely separated volcanoes, one trending NE.–SW., and the other NW.–SE. (See Fig. 4). One of these, the well known "Cameroon fissure," bisects the Gulf of Guinea and includes the volcanic islands, St. Helena, Annobom, Sao Thomé, I. do Principe, and Fernando Po. On the land this fissure is continued in a striking manner by the fault bridge which ends in the Tschebitschi, 2,000 meters high, which then drops suddenly to the level of a low plain less than 200 meters above the sea. The volcanotectonic line which intersects this striking lineament at St. Helena, includes Ascension, one of the eastern cones of St. Paul's Rocks and a conical, submerged elevation upon the sea floor, almost under the tropic of Capricorn about 800 kilometers southwest of Amboland.

In addition to these two fissure directions, a third is like them strikingly characteristic of the African continent, as shown by the remarkable north and south lines of volcanoes and rift valleys in central Africa east of the Nile. To these three prevailing directions, northwest-southeast, northeast-southwest, and north-south, must be added a fourth less common direction, namely, east-west. Simmer

[29] Hobbs, *Gerlands Beiträge z. Geophysik*, Vol. 8, 1907, pp. 316–317.

[30] *Ibid.*, pp. 315–316, smaller map of pl. 3. See also Suess, "The Face of the Earth," Vol. 1, p. 144.

[31] Hobbs, *l. c.*, pp. 348–349, pl. 10.

in a noteworthy compilation[32] has shown that these directions are brought out for the African continent not only in the lines of

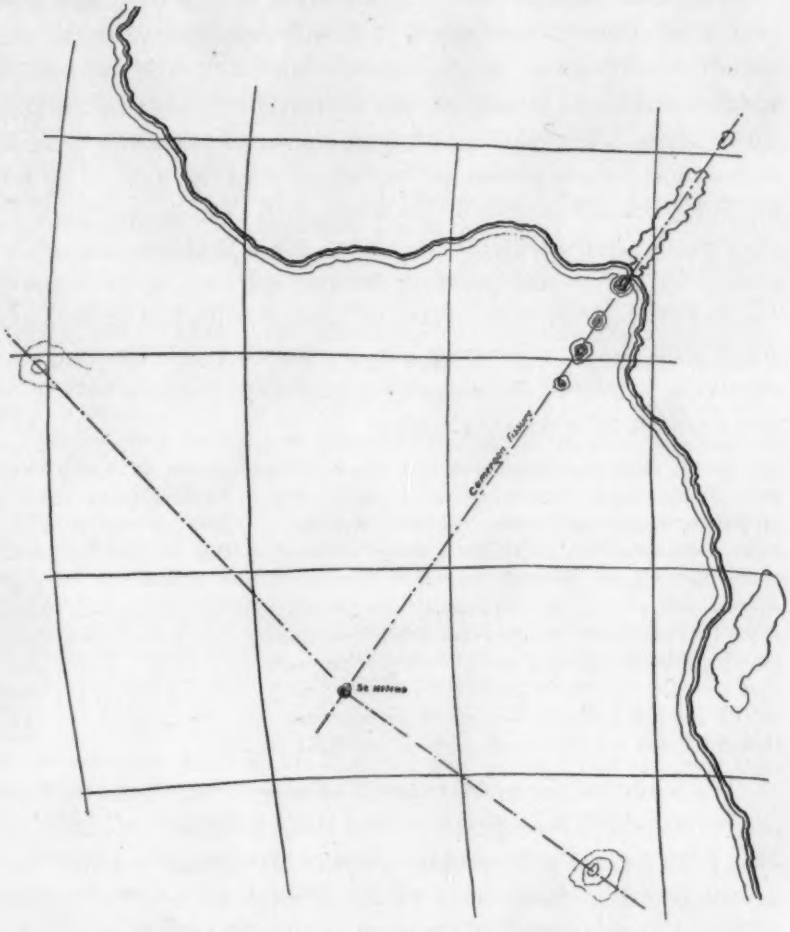


FIG. 4. Volcanotectonic lines which cross at St. Helena.

volcanoes but by the fracture systems revealed in the rocks, so far as they have been studied. It is interesting to note that these

[32] "Der aktive Vulcanismus auf dem Afrikanischen Festlande und den Afrikanischen Inseln," *Münchener Geographische Studien*, No. 18, 1906, 218 pp.

directions are also the dominant ones in the fracture system of North America.[33]

A number of papers of a controversial nature have appeared notably by Branca[34] in opposition to the view that volcanoes are aligned upon fissures, but inasmuch as they deal with districts in which the evidence is more or less equivocal they need not be considered here. The problem of arrangement of volcanoes must be solved not in southern Europe nor on the Mexican plateau border, but in the volcano gardens of the world, such as Iceland or Java.

Volcanic Extrusions in Relation to Block Adjustments.—Clarence King in his description of the area included in the Fortieth Parallel Survey,[35] an area divided by vertical faults into great blocks which underwent adjustments at the close of the Miocene, has furnished a classical instance of the relation of volcanic outflow of lava to block movement. He says:

> Single ranges were divided into three or four blocks, of which some sank thousands of feet below the level of others. The greatest rhyolite eruptions accompanied these loci of subsidence. Where a great mountain block has been detached from its direct connections and dropped below the surrounding levels, there the rhyolites have overflowed it and built up great accumulations of ejecta. Whenever the rhyolites, on the other hand, accompany the relatively elevated mountain-blocks, they are present merely as bordering bands skirting the foothills of the mountain mass. There are a few instances in which hill masses were riven by dykes from which there was a limited outflow over the high summits—but the general law was, that the great ejections took place in subsided regions.

The study of the great rifts of eastern Africa seems to have shown that the volcanoes which have there been built up, are similarly related to the sinking of the great strips of country which have caused the chief inequalities of the general surface.[36] The two

[33] Hobbs, "The Correlation of Fracture Systems and the Evidences of Planetary Dislocations within the Earth's Crust," *Trans. Wis. Acad. Sci.*, Vol. 15, 1905, pp. 15–29.

[34] W. Branca, "Zur Spaltenfrage der Vulkane," *Sitzungsber. Ak. Wiss.*, Berlin, 1903, pp. 748–756.

[35] "United States Exploration of the Fortieth Parallel," Vol. 1, Systematic Geology, 1878, p. 694.

[36] Ed. Suess, "Die Brüche des östlichen Afrika," *Denksch. Weiner Akad., Math. Naturw. Kl.*, Vol. 58, 1891, pp. 555–584.

chains of volcanoes in Mexico as mapped by Sapper[37] seem to be similarly associated with the great rift valley lying on the western border of the Mexican plateau.

It is in Iceland, however, that the most extended studies have been made of the most interesting field, in which the relation has been worked out with the greatest thoroughness.[38] Says Thoroddsen:

> One gains the impression that the form of the surface has no significance as regards the volcanic force, which breaks out above upon the ridges, as well as below in the valley, yet the volcanoes are always found associated with areas which are either sinking or have sunk.
>
> The lava stream Ögmundarhraun in Krisnoik, which dates from about 1340, was poured out from two parallel clefts. The southernmost portion of this stretch of country between the clefts after the beginning of the eruption sank about 66 meters, and one side of the western fissure rose like a vertical wall with four half craters open at the brink, the other halves having sunk. At the end of the cleft is a visible dike which leads up to the row of craters.
>
> Where great fractures or faults are present in the crust, the volcanic forces have not always made a single passageway through them, but in the vicinity on parallel clefts, often upon the high fracture margin; thus one fracture line 50 km. long extends without volcanoes from Krisnoik to Hengill, at which place the north side is sunk 200 to 300 meters; parallel with this is here found above at the margin of the cliff an almost uninterrupted series of craters which have formed not alone upon a single fissure but over several slices and small fissures running parallel to one another. A similar phenomenon is to be observed on the southern fracture margin of the peninsula of Snaefellsnes where the craters are mainly found above upon the edge of the bluff. Often, also, the reverse is the case, as for example, in the Odadahraun, where the rows of craters for the most part extend along the bases of the mountain chains, which rise as horsts from the sunken ground on either side; a like example occurs at Myvatu, although here the rows of craters occur at times above upon the ridge.

In none of these cases have we evidence that the eruptions coincided closely in time with the earthquakes which must have accompanied the movements of the earth strips between their bounding faults, but the relationship of the one phenomenon to the other could hardly be more clearly proven. Summing up the discussion, we note that volcanoes, no less than earthquakes, help us to find the positions of those fissures within the crust by which it is separated

[37] "Ueber die raümliche Anordnung der Mexikanischen Vulkane," *Zeitsch. d. Deutsch. Geol. Gesell.*, 1893, pp. 574–577.

[38] *L. c.*, p. 3.

into a mosaic of blocks, and that these lines of fracture may therefore be designated *seismotectonic* or *volcanotectonic lines* or simply *lineaments* according as they are revealed by earthquakes, by volcano rows, or by topographic and geologic peculiarities.

A Possible Explanation of "Volcanic Earthquakes."—Writing before 1885 Suess distinguished two classes of earthquakes, the dislocation and the volcanic earthquakes, and to these Rudolph Hoernes added the type of in-caving earthquakes to cover especially some of the light shocks of the Dalmatian coast. If we were to supply a complete category of earthquakes it would be necessary to add further a type of cataract earthquakes to cover the occasional fall of limestone blocks in the Niagara cataract, as well as many other minor forms, such as blast shocks in mines, etc. In point of importance two classes only stand out sharply as they were originally announced by Suess, and the present writer has been of the opinion that even these may perhaps be subclasses only of a single phenomenon. The mechanics of volcanic eruption, so far as it applies to the cone, is now so well understood that we are able to connect the outflow of lava which marks the beginning of the grand stage of paroxysmal eruption in a composite cone, with the rending of the mountain and the opening of a fissure—a distinctly tectonic movement induced by the lava as it rises under the influence of gravity, aided perhaps by the expansive power of the associated steam. I believe we have been misled into supposing that the fissures which are thus opened are necessarily radial to the cone, since this would be presumed if the mass of the cone and its basement were throughout homogenous, with no preëxisting fractures, and were acted upon by hydrostatic pressure from the central shaft only.

Etna is a giant mountain rising nearly 11,000 feet directly from the sea, its diameter is more than twenty-five miles, and since the higher portions are so largely concentrated at the center, the average thickness of visible volcanic ejectamenta over the base of the cone is only about one half mile. Apparently, therefore, this superficial layer of volcanic material may play a relatively small rôle in the rending of the entire mass which accompanies an outflow of lava. So soon as we examine the lines of parasitic craters which

are distributed upon the flanks of the mountain, we find that the majority of these are not radial to the mass at all, but comprise a network. A notable instance of a line of craters not in radial relation to the central cone is furnished by the chain of Monti Segreta, Nocella, Pizzuta, Gervasi, Arso and Difeso. Nearly parallel to this chain is that of the Monti Mazzo, S. Leo, Rinazzi, Guardiola and Albano. A map of these and other monticules upon the flanks of Etna has been already published by the writer.[39] It is, therefore, not only possible, but extremely probable, that in many instances the earthquakes which so generally accompany the rending of a volcanic cone, are directly associated with the opening of, and perhaps a differential movement upon, those fractures in the basement of the mountain which are a part of the larger fracture system of the district. Lacroix has recently shown that a *network* of fissures appeared upon Etna in connection with the eruption of 1908.[39a]

The Conditions of Earth Strain During the Growth of Block Mountains.—If we consider any circumscribed portion of the earth's crust within which mountains are growing through the adjustment by individual blocks or compartments of the crust, it is necessary to assume that the superficies is increased during the process. Individual blocks may indeed be actually depressed as a consequence of the adjustment, but yet the average movement must be assumed to be upward rather than downward. Such a conclusion is, however, in contradiction of the generally accepted view that mountain growth comes about through a reduction of superficial area from secular cooling. This very obvious difficulty in the way of adopting the *Schollen* conception of mountain structure has been quite generally recognized, and we have already seen how Oldham, in seeking the cause of the great Assam earthquake, was led to reject the theory, even though the vertical faults and the differential changes in level were plainly to be observed.

In the opinion of the writer, the recent study of "distant" earthquakes by modern seismographs has removed this difficulty in the way of a general acceptance of the fault-block theory. By extend-

[39] *Gerland's Beitraege z. Geophysik*, Vol. 8, 1907, pp. 348–350, Pl. 10.

[39a] L'eruption de l'Etna en avril-mai 1908, *Revue générale des Sciences pures et appliquées*. 20e année, 1909, pp. 298–314.

ing our knowledge of surface displacements of the earth to the floor of the oceans, it has brought us a surprise; for we have learned that to these areas, by many regarded as so stable, belong a much larger proportion of the grander movements, and by presumption of the smaller ones as well. The recent study of the ocean floor through soundings, examined with reference to the loci of suboceanic quakes, has told us, further, that though the movements upon the land are generally upward, those upon the ocean bottom, on the contrary, are downward. The so-called "origins" of the oceanic quakings are most frequently the steep borders of the great sea troughs where the greatest depths have been revealed by soundings. Now it is as impossible to separate the idea of molar displacements from these great disturbances as it is to avoid the conclusion that since these troughs are now the deepest bottoms, this is a direct consequence of the repeated displacements which must accompany the quakings. It has, moreover, been a general result of direct observation, that with noteworthy local exceptions the sea-coasts are to-day undergoing elevation, and that the steeper coasts face the greater depths.[40]

It is difficult to avoid the conclusion that the general upward movement of the margins of the continental areas and the general downward movements of the near-lying oceanic floors are interrelated as parts of one general adjustment within the outer shell of our planet. This granted, there is no difficulty in conceiving of the rise of block mountains upon the continental borders, since the increase of superficies within the affected continental region is compensated by a contraction of area in portions of the sea floor which in the same general period are subsiding. A rise of block mountains to the accompaniment of an earthquake, if our theory of cause be correct, though it calls for an expansion of the surface, should reduce the superficies of the affected region *if measured on the surface of a sphere at its former level.* A renewed and sudden compression of the district is thus made possible through the action of the tangential compressive stresses within the contracting shell. The writer believes that evidence of such compression has been

[40] See, among others, G. Schott u. P. Perlewitz, "Lothungen I. N. M. S. "Edi" und des Kabeldampfers "Stephan" im westlichen Stillen Ozean," *Arch. d. deutsch Seewarte,* Vol. 29, 1906, pp. 5–11.

found in the case of most large earthquakes in the behavior of rails and bridges.[41]

Part II: The Outlook of Seismic Geology.

The Ultimate Cause of Earthquakes.—No one should be deceived into concluding that because we seem to have found some evidence of the nature of the process by which the external shell of our planet undergoes its adjustment at the time of an earth shock, we have thereby discovered the *ultimate* cause of earthquakes. That is a far deeper problem, to which the discovery of the *proximate* cause is but an initial stepping stone. It is in this field that the deeper secrets lie hidden. The outlook of the science indicates two lines of effort to be followed up. These are: (1) To make practical application of the knowledge already gained, and (2) to investigate with every possible improvement in method until we have so laid bare the laws of seisms that we may forecast the time, the place and the probable severity of future earthquakes with at least as much accuracy and forewarning as is now possible in weather prediction.

Earthquake Forecasts.—It is much to be feared that the science of earthquakes is to pass through a stage not unlike that in meteorology which ushered in the day of scientific prognostication. Judging from statements which have been published, a "Farmer's Almanac" of earthquakes and popular earthquake prophets may be looked for as a possibility of the near future. It will be well, therefore, to consider the nature of the earthquake forecasts which have been so widely advertised. Examined with care it is found that these, in so far as they have found any verification, apply to a single, though the most important, seismic zone, and that all are indefinite as to the time and largely so as to place. Dr. Omori, of Tokyo, after the California earthquake of 1906, made a forecast which he himself subsequent to its partial verification reported as follows:[42]

> As to the probable position of the next great shock on the Pacific side of America I expressed my view that it would be to the south of the equator

[41] Hobbs, "A Study of the Damage to Bridges During Earthquakes," *Jour. Geol.*, Vol. 16, 1908, pp. 636–653.

[42] *Bull. E. I. C.*, Vol. 1, No. 1, p. 23.

(that is to say, Chili and Peru), as it was very likely that the seismic activity would extend to either end along the great zone in question, and as the coasts of the countries above named are often visited by strong earth convulsions.

About two months after the prediction was made occurred the Valparaiso earthquake, but at the same hour an earthquake of the same order of magnitude visited an area in the Aleutian Islands within the same seismic belt, though nearer and in the opposite direction from the one predicted. On the same grounds Lawson in a lecture read in March, 1907, said of the stretches between southern California and Central America, and between northern California and southern Alaska:

These strips, I believe, will be visited before long, and then the long line of this earthquake will be complete from Chili to Alaska.

The Guerrero earthquake in Mexico occurred only a few weeks later and bore out the geologist's faith in the soundness of his hypothesis.

The method upon which such predictions are based is already indicated in the quotations given. Briefly expressed it is the principle of immunity from shock for a considerable period after heavy earthquakes, combined with the conception of relief secured throughout an extended zone in sections by alternation. An extended zone on the earth's surface is recognized to be what might be called an orographic unit; that is to say, it is all undergoing progressive though interrupted elevation. Stresses tending to produce uplift are presumably cumulative and may be of varying amounts in different sections of the zone. The resistance to movement under the strain—whether due to the rigidity, to the vice-like compression, to the absence of suitable fissure planes on which the movement might occur, to the healing of such fissures by mineral matter, or to any other causes—may be assumed to be different in different parts of the zone. Relief of stress through sudden uplift should, therefore, occur first within some one section of the zone where stresses are greatest, resistance least, or both. The earthquakes furnish abundant proof of the general correctness of this view. Now it is simpler to assume that relief having been secured in one section of the belt, a certain lowering of the potential energy of the system of

stresses is to be expected in the near-lying sections on either side, particularly since the shock tends to discharge the system of strain as would a fulminate. On the theory of probabilities the area next to be relieved should be the most distant, providing stress has there been accumulated for an equally long period. The third and fourth steps in the cycle of release of strain should in position be intermediate between the first and second on one side or the other. Later steps in the "letting down" process should affect especially the still intermediate unrelieved sections of the zone.

This method, simple as it is in theory, permits of only the broadest generalization and, as already stated, has been tested in but one zone and for one cycle of relief. This zone is the great circle belt which surrounds the Pacific Ocean, and the cycle of relief seems to have begun with the Colombian earthquake of January, 1906. Only two months after this disturbance came the Formosa earthquake, in a province between one third and one half the distance around the planet. The area of the California earthquake, which occurred a month after that in Formosa, is intermediate between the first two, though nearer the first than the second. By examination of Fig. 5, which is drawn to scale, it will be noted that the distances separating the approximate centers of these and the later disturbances in the series, generally bear out the hypothesis that each later earthquake affects an area farthest removed from those sections of the zone which have already found relief.

The rapidity with which the steps in the process of securing relief have here succeeded to one another, lends strong support to the view that the zone in question should be regarded as a definite orographic unit, and that the stress-strain conditions within all except the southernmost portions were before relief began, remarkably uniform. The planetary order of magnitude of the movements would thus seem to be clearly indicated. The section of the zone last to be relieved was, it is interesting to note, one which had been partly relieved of stress during two earthquakes six years and four years before the main cycle of relief was inaugurated. The section which separates the district of the Aleutian from that of the Californian earthquake had also been visited by earthquakes seven years and six years previous to the main cycle of relief. The portions

of the zone in which the probability of heavy shocks is now most imminent, are the Japan-Kamschatka segment, the Peru-Bolivian segment, and the archipelago region to the southeast of Asia. Inasmuch, however, as between 1899 and 1903, 29, 12 and 41 heavy shocks had been registered by seismographs from the vicinity of

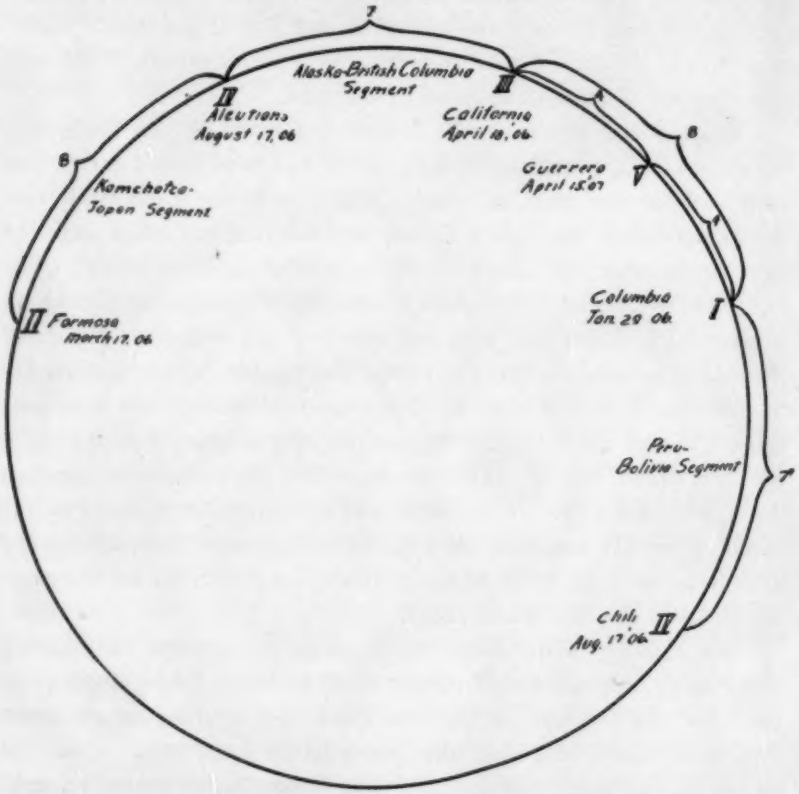


FIG. 5. Diagram showing the distances which separated the approximate centers of areas of the series of earthquakes within the circum-Pacific zone in the years 1906–7.

these three segments respectively,[42] the time may be long before the limit of strain may again be reached in them. The problem is thus far from simple and prediction would be extremely hazardous.

It should not be forgotten that prediction of any sort has thus

[42] Milne, *Geogr. Jour.*, 1903, map.

far been possible only within this circum-Pacific zone, which, at the time, is passing through a remarkable seismic history. It is little likely that any such sudden relief of strain will take place again in the same zone before a considerable period has elapsed.

Yet, outside this zone and within our own country, earthquakes of the first order of magnitude have visited the lower Mississippi Valley, the coastal plain in South Carolina and the valley of the St. Lawrence during the brief period that the country has been occupied by whites. Of these sections of country, as of most others, the only safe prediction that can now be made, is that districts already visited by historical destructive shocks, as well as some others, notably New England and the Middle States, will eventually suffer from disastrous earthquakes. To the time of such visitations we have not even a clue.

Periodicity of Earthquake Cycles.—The "letting down" of the potential energy of the system of stresses within the circum-Pacific belt, as brought out by the events of 1906–7, is, in the writer's belief, as regards its close sequence, an event without parallel in the history of seismic geology. Something approaching it appears, however, to have been in operation within a somewhat longer period in the other great seismic belt of the globe. Making all due allowance for the fact that our quite recent study of distant earthquakes has greatly extended our horizon, it still seems necessary to conclude that the present is a time of very exceptional seismic intensity.

So soon as we admit the planetary scale of these seismic disturbances and explain them as a result of mountain growth upon the borders of the continent, we are led to expect the existence of such maxima and minima of seismic intensity. If now we examine the history of earthquakes in those countries possessing the longest records, we find evidence in support of this view. The stronger earthquakes in Japan, which are on record for a period of fifteen hundred years, betray a strong tendency to group themselves. The 154 heavy earthquakes recorded in that country since the beginning of the fourteenth century may be divided more or less definitely into 41 groups separated by average intervals of 13½ years. In Kyoto a complete record has been kept for a thousand years. Here there was a strong maximum of destructive and strong earthquakes be-

tween the middle of the fourteenth and the middle of the fifteenth century, this maximum period being followed by a steady decrease to a minimum in the last half of the nineteenth century. Minor fluctuations reveal an average period of 6½ years, or about one half that revealed by the records for the Empire as a whole.[44]

The natural objection which would be raised to making use of these data for basing conclusions upon the behavior of the earth as a whole, is that the maximum of intensity in Japan may well have been compensated by a minimum in a neighboring district. What we need for basing our conclusions is a world catalogue of earthquakes extending over a sufficiently extended period. Thanks to John Milne and those who have followed his lead, we are now preparing such a catalogue, which is sure to permit of a definitive answer to the question of earthquake periodicity. Even within the first section of this catalogue, comprising as it does the thirteen years from 1892 to 1904, Milne believes he has made out a relatively short period with the maxima of world shaking in correspondence with the more abrupt changes in direction in the orbit of the earth's pole. On *a priori* grounds it is reasonable to connect seismic disturbances with sudden changes in latitude, and the further data upon the pole movement and the seismic world maxima, will be scrutinized with interest.

Possibilities of Future Prognostication.—It is too early to predict whether more satisfactory bases for future forecasting of earthquakes will be discovered, but the indications are certainly encouraging. Two, and perhaps three, lines of inquiry are already suggested. Most promising of these, is, perhaps, the study of terrestrial magnetism; for in a considerable number of instances, destructive earthquakes have been preceded by periods measured in hours and sometimes in days, within which the behavior of magnetographs was singularly abnormal. It seems likely that this change in magnetic conditions may sometimes be utilized as a warning signal. For solution of this problem the completion of the magnetic survey of the world, may be expected to contribute.

Evidence is not lacking that fore-shocks, or rather fore-tremors,

[44] Kichuchi, E. I. C. Pub., No. 19, 1904, pp. 11–13.

for they would appear to have an extremely small amplitude of vibration, are a fore-runner of most heavy earthquakes. These fore-tremors should not be confused with the preliminary tremors in the record of the distant seismograph, for they are of such small amplitude that they would probably not be registered by any instruments today constructed, except perhaps within the affected district itself. Our best evidence that such fore-tremors exist is furnished by the behavior of certain of the lower animals. In the opinion of the writer, such a body of evidence has now accumulated, that it can no longer be waved aside. Just as the sense of smell is so much more highly developed in the dog, for example, than it is in man, so there seems no valid reason for doubting that the detection of small motions by the lower animals may be by as much superior to the human sensibility. Dr. Omori has expressed his belief that seismographs will yet be made sufficiently sensitive to record these microscopic tremors. Just as a block tested in our experiments assumes very large deformations as it approaches rupture, so the earth structure may behave during a period which is as much longer in proportion as the time of augmenting the stresses exceeds that in our experiments. Judging from the recorded behavior of animals, it would not be surprising if the period during which warning may be possible on this basis, should prove to be a large fraction of a day, or even longer. If measurable deformation does occur as a result of the accumulated stresses long before the limit is reached, it may be possible in the case of those earthquakes particularly which result in horizontal shearing movements, to determine by frequent measurement of the distances which separate properly placed monuments, the approach of the strain limit. It is a subject which is at least worthy of investigation.

Since the days of Perrey, who devoted his life to an attempt to find a connection between earthquakes and lunar conditions, there have been those who have sought to connect seismic and volcanic disturbances with periods of special gravitational stress due to luni-solar phases. The most recent advocate of such a connection, is Perret,[45] who is so convinced that he has found the secret behind

[45] "Some Conditions Affecting Volcanic Eruptions," *Science*, Vol. 29, 1908, pp. 277–287.

the phenomena as to have ventured to predict for the year 1908, a grand eruption of Etna.[46] This eruption not having materialized, Perret has accepted the Messina earthquake as a substitute.[47] Assuming that his method is correct, it is possible to see how a period of seismic or volcanic activity might be predicted; the method, however, gives no clue as to what part of the earth's surface is likely to be thus affected. The predictions of the author of the theory have,

FIG. 6. Abandoned Sea Cave 10 feet above water on Coast of California. (After Fairbanks.)

on the whole, been less remarkable than the statements made by one of his supporters.[48]

Need of an Expeditionary Corps.—It may well occasion surprise that governments have been so slow to appreciate the necessity for providing means for the investigation of earthquakes. Our own government, which has shown such commendable generosity in providing the sinews for scientific investigation, has in this particular

[46] *The World's Work*, November, 1907.
[47] *Am. Jour. Sci.*, Vol. 27, 1909, pp. 322–323.
[48] Jaggar, *The Nation*, Vol. 88, 1909, pp. 22–23.

field lagged far behind other nations. In Japan since 1892 there has been an Earthquake Investigation Committee, and whenever a destructive earthquake is reported from any part of the world, Professor Omori, the secretary of the committee and its chief expert, is despatched by his government to prepare a report upon it. Under orders from the Japanese government, he is today in the vicinity of Messina engaged in a study of the latest great disaster. While these expeditions have been of value in securing information, the time has come when with the incease of our knowledge of earthquakes, something more than a reconnaissance survey is required. One man without assistants and without elaborate equipment, is today in no position to secure those more important data which alone can advance our knowledge of earthquakes beyond its present status. Today a scientific party should have at its disposal one or more surveying vessels—small gunboats or protected cruisers could be easily adapted for the purpose—provided with modern sounding apparatus and with a full equipment of necessary instruments. The crops of scientific workers should include skillful topographers and their assistants and all suitable instruments for preparing accurate topographic maps. The party should also include trained experts whose duty it should be, among other things, to map the distribution of the surface intensity of the shocks. An expeditionary vessel of the type described could be utilized upon occasion to study volcanic as well as seismic disturbances; such, for example, as the late eruptions in the Windward Islands. The seismic events of the years 1906–8, would have been more than sufficient to take up the attention of two surveying vessels with their corps of scientific workers.[49] In times of relative seismic inactivity the ships and their complements could be employed to advantage in work which will be more definitely indicated below.

A Service of Correlated Earthquake Observatories.—In addition to the study upon the ground, which may be expected to lay bare some important laws of seismic geology, there should be installed a series of stations equipped with modern seismographs for the regis-

[49] In a late number of the *Popular Science Monthly* (February, 1909) the writer has pointed out the exceptional opportunities which the recent Messina disaster has offered for study by this method.

tration of the distant as well as the nearer and local earthquakes. These stations should be well distributed over the national domain, and should include a number of stations of the first rank provided with the more sensitive type of pendulum adapted to the registration of distant earthquakes. A larger number of stations of lower rank should be provided with simpler instruments suited only for securing full data upon the local shocks. These smaller stations should be located with due regard to the more important seismic provinces of the country. The United States Weather Bureau already possesses suitable buildings for installing such apparatus, and the regular employees of the stations could be trained to add the care of the instruments to their other duties. In 1907 with the hearty approval of the heads of the various scientific bureaus of the government, the American Association for the Advancement of Science, upon recommendation of its Committee on Seismology, memorialized Congress upon the pressing need of such a service. A year later, the Geological Society of America passed a resolution of similar import, and in the same year, no positive result having been secured, the Committee on Seismology renewed its first memorial by a second resolution.[49a]

Scientific research has already gone far to remove some of the greatest scourges of human existence. Of those which are characterized by sudden and usually unexpected visitation, are pestilence, flood, conflagration, earthquake and volcanic eruption. Of these flood and conflagration must be in part laid at the door of earthquake disturbances, to which they have all too frequently been an almost inevitable sequel. They have, moreover, taken the larger toll of human life and property. As compared with epidemic diseases, like the plague and smallpox which repeatedly overran Europe during the middle ages, earthquakes and their consequences have been the less destructive of life. It has been estimated that in Europe, the plague alone carried off no less than 25,000,000 people. Yet medical science has discovered the mystery of the disease, and in sanitation and isolation provided the remedy. To meet the great dangers of conflagrations, which from time to time

[49a] See also the resolution passed by the American Philosophical Society on April 24, 1909. *Proceedings* No. 191, p. xii.

have swept over our cities, we have as yet made only partial provision, yet the remedy is known and the country does not hesitate to make an annual expenditure conservatively estimated at $25,000,000, and in addition compels its citizens to build according to approved regulations.

A single earthquake has involved us in a loss of over $350,000,000, or nearly ten times the loss from the Baltimore fire.[50] Yet the government has expended nothing in an attempt to safeguard the future by avoiding the recurrence of such disasters. In Europe within a few months an entire city has been laid in ruins with a loss of life which may reach 150,000, yet the latest information makes it almost certain that this quake was not an exceptionally heavy one, and that most of the loss of life and property might have been avoided if proper methods of construction had been adopted.

It can hardly be claimed that the comparatively recent California disaster gave us our first warning of danger, for twenty years earlier the earthquake in South Carolina caused a loss of over one hundred lives, and property to the value of between $5,000,000 and $6,000,000. The earlier earthquakes within our territory have been far heavier and the small loss of life and property is accounted for only because the districts were at the time so thinly populated. We must not, therefore, overlook the fact that the United States is an earthquake country, and this not alone in its Pacific section. Some of our largest and most prosperous cities are almost certain to pass through their trials in the future, as Charleston and San Francisco have so recently. On February 5, 1663, almost the entire valley of the St. Lawrence and large sections of New England were visited by an earthquake, which, if the country had been built up as it is today, would have caused a disaster which it is not pleasant to contemplate.

Preparation of Maps of Fracture Systems.—As we have seen, earthquakes register the movement of portions of the earth's crust between planes of fracture. In just how far these fracture planes are present in advance of the movement, and in how far they result

[50] The official figures kindly furnished by Professor J. W. Glover.

from the relief of strain at the time of the shocks, has not yet been determined. Some writers have dismissed from consideration as "secondary phenomena" most of those visible fractures which first appear at the surface during an earthquake. It seems certain, however, that many of these fractures, at least, as regards both direction and position, are dependent upon the fracture system already present in the underlying rocks; and there is, therefore, need for extended study of the fracture and fault system within the rock basement of each earthquake province. With this study might perhaps be combined the determination of the depth and the earthquake properties of each of the overlying unconsolidated deposits. Experiments are further necessary in order to determine whether large thicknesses of such deposits are controlled by the same laws as are the thinner ones.

In every district which has an earthquake history, this record should be examined to learn if possible the points, the lines, or the areas of heaviest shock. Whenever data are sufficiently complete, maps should be compared to represent the approximate distribution of surface intensity for each earthquake, and comparisons instituted.

Maps of Visible Faults and Fissures and of Block Movements for Special Earthquakes.—It has been pointed out that in the case of a single earthquake only has a map been prepared to show in detail the distribution of the surface faults and the block movements of the ground. Thirty-five years after the event which brought them into existence, these faults have been mapped in detail by Mr. W. D. Johnson, of the United States Geological Survey. It has been possible to prepare maps of portions only of the district affected, and the full results are not yet published. Within the national damain there are at least two other provinces which promise fruitful results from such a study. These are the regions affected by the Sonora earthquake of 1887, and, even more important, the country about Yakutat Bay, Alaska, so profoundly modified in its relief during the earthquakes of 1899. A scientific party with headquarters upon a surveying vessel, such as we have described, would here find almost unequaled opportunities for securing important data.

Rate of Mountain or Shore Elevation by Quantitative Methods.

Wave-cut terrace and sea arch 10 feet above present sea level on coast of California. (After Fairbanks.)

Elevated shore line exhibiting stacks on coast of California near Port Harford. (After Fairbanks.)

—The studies of earthquakes during the last few years have done much to destroy the illusions of more than half a century. Since the time of Lyell, the burden of all geological instruction has been the extreme slowness of terrestrial dynamic processes. Oscillations of level described as slow and uniform warpings of the crust, had been gauged by measurement upon shores, which in the expressive language of de Montessus are dead, and where in consequence earthquakes are seldom or never left. If movements accomplished within a week and largely upon a single day, can elevate stretches of coast over 47 feet, as was true of portions of the Alaskan coast in 1899, what modifications of our traditional theories will be required! There is a pressing need for extended studies on rising coasts to determine by some scale the rate of elevation.

Now it happens that one of the most rapid of erosional processes is that accomplished by the waves as they beat upon a lee shore, and this process is one capable of fairly accurate quantitative measurement. The Pacific coast of North and South America, the greater part of the way from Alaska to Patagonia, has, during a recent period, been rising to the accompaniment of earth shocks. As we now understand, these uplifts have been mainly spasmodic, and the strand-lines abandoned with each successive uplift now stand revealed in a series of steps or terraces, which, when closely examined, reveal the characteristic marks of wave action sometimes at heights of fifteen hundred feet and more (see Fig. 7, and Plates XV. and XVI.). Careful maps prepared after correlation of these strand lines throughout long distances when combined with precise studies of the rate of wave cutting, could hardly fail to shed light upon the broader problems of seismic geology.

In some cases such abandoned shores now in an elevated position reveal clearly that their uplift was sudden and that no interval long enough to permit wave cutting separated it from the inauguration of the present level. Thus in figures 6 and 8 are represented shores which might almost be described as fossilized earthquakes, for the evidence is clear that the elevation took place in what was essentially a single sudden stage and must have been accompanied by a great quake.

Seacoasts offer the best possible data for observation and meas-

urement of the rate of uplift, because the level of the water can be made use of for the zero point. There are, however, other available means for investigating the rate of continental uplift. In arid and semi-arid regions, such as the Great Basin of the United States, the

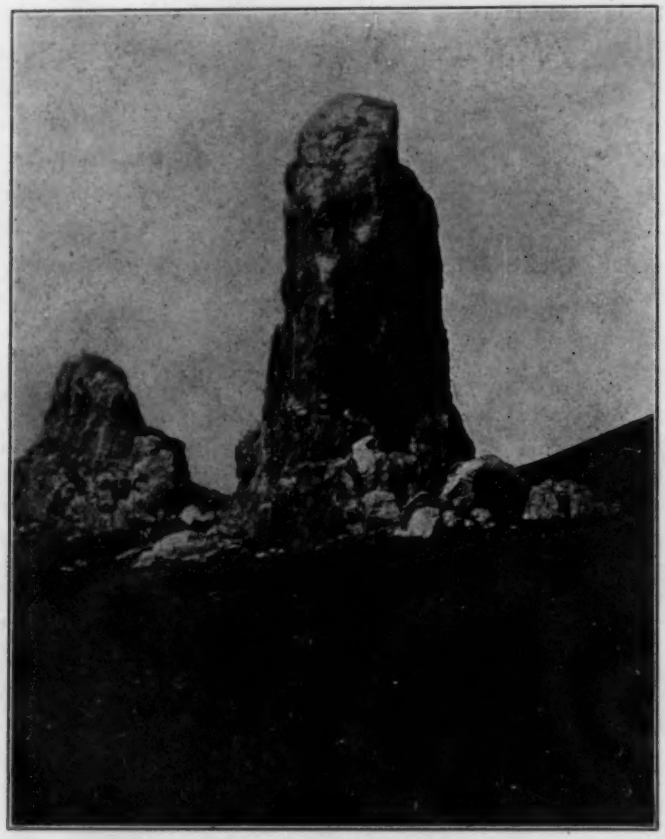


FIG. 7. Elevated shore on the coast of California showing marks of wave action. (After Fairbanks.)

rare but violent storms cause torrents in the streams which debouch upon the plains from the mountain fronts, and so broad fans and aprons are there built up. Now if the uplift of the range goes on more slowly than the alluviation along its borders, the mountain front deposits will bury and hide the escarpments which are opened at the time of each successive uplift. If, on the other hand, the

uplift is the more rapid, fault scarps will appear cutting the unconsolidated deposits. Such scarps, some of them twenty feet in height, are characteristic of both the Eastern and the Western margins of

FIG. 8. Elevated and present shore lines registered in notches of chalk cliff at Cape Ciro, Celebes. (After Paul and Fritz Sarasin.)

the Great Basin region. From careful study of the rate of deposition there is here the possibility of reaching an approximate measure of the rate of uplift.

Investigation of Earthquake Water Waves.—The great water wave which followed the famous Lisbon earthquake of 1755 was

more destructive to human life than the shocks which proceeded it. The earthquake water wave which inundated the shore of Japan on June 15, 1896, destroyed human lives to the number of 29,953. Such waves have been especially destructive along the western coast of South America. The new seismology, by instrumental methods, points more and more definitely to the cause of such disturbances in the subsidence of great sections of the neighboring ocean floor; yet with the exception of relatively small waves within the Mediterranean, we are without observational data in the form of soundings in confirmation of this hypothesis. The bottom of the ocean is each year being charted in new areas, and we are fast accumulating data on which to base a decisive series of observations to settle this important question. This will certainly be one of the larger problems for investigation in seismic geology.

Conclusion.—It has been possible to indicate a few only of those directions along which effort will be directed in the early future of seismic geology. From this summary, I think it will be seen that there remain no other fields of investigation so long neglected and yet so full of promise in important discoveries, which are likely to touch so intimately the lives and happiness of human beings. What we have already learned is much of it as yet only half learned, and we need careful experimentation on lines already marked out, so that recommendations may be made for adapting our lives to future seismic conditions. Probably nine tenths of the danger from earthquakes can be avoided through practical methods of construction, but the relative cost of the different means of securing immunity must be carefully considered. The studies which are necessary are on such a scale that they call for generous government support, and it cannot be too strongly urged that the United States government undertake a work so clearly demanded by the situation. This support should be nothing less than the foundation of a bureau for earthquake investigation, with regular appropriations sufficient to carry out studies by a system of correlated earthquake stations, and also upon the ground of each devastated region whether it be at home or abroad.

University of Michigan,
April 21, 1909.

SEISMOLOGICAL NOTES.

By HARRY FIELDING REID.

(*Read April 24, 1909.*)

(*a*) Conditions Preceding and Leading to Tectonic Earthquakes.

There are two classes of earthquakes: Volcanic and Tectonic; the former, connected with volcanic outbursts, seem to be due to explosions or to the sudden liberation of steam; the latter are due to ruptures of the rock. It is only the latter class that we shall consider at present.

Rock, like all solids, is elastic, and when subjected to external forces it suffers an elastic strain; if this strain is too great for the strength of the rock to withstand a rupture occurs; but it is never possible for a rupture to take place until the rock has been deformed or stretched beyond its elastic limit. When the rupture occurs, the two sides spring apart under the elastic forces and come to positions of equilibrium, free of elastic strains. The following experiments have been made to illustrate these conditions. Two short pieces of wood were connected by a sheet of stiff jelly 1 cm. thick, 4 cm. wide and about 6 cm. long, as shown in Fig. 1. The jelly was cut through along the line, tt', by a sharp knife and a straight line, AC, was drawn in ink on its surface. The left piece of wood was then shifted about 1 cm. in the direction of t', and a gentle pressure was applied to prevent the jelly from slipping on the cut surface. The jelly was sheared elastically and the line took the position AC shown in Fig. 2. On relieving the pressure so that the friction was no longer sufficient to keep the jelly strained, the two sides slipped along the surface tt' and the line AC broke into the two parts AE and DC. At the time of the slip A and C remained stationary, and the amount of the slip, DE, equalled the shift which A had originally experienced. A straight line, $A'C'$, was drawn on the jelly after the left side had been shifted, but before the jelly slipped along tt'. At the

time of the slip, the same movement took place in the neighborhood of this line, as near AC, and $A'C'$ was broken into two parts, $A'E'$ and $D'C'$; the total slip, $D'E'$, being equal to DE. A third experiment was tried; the left piece of wood was shifted 1 cm. and a straight line was drawn across it; it was then shifted a half centimeter more and the straight line took the position $A''C''$ in Fig. 3. When the jelly slipped along the surface, tt', the line broke into the

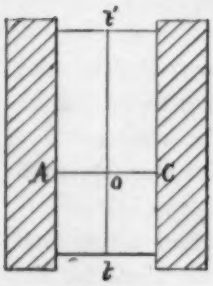


Fig. 1.

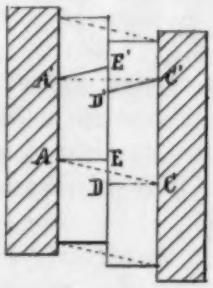


Fig. 2.

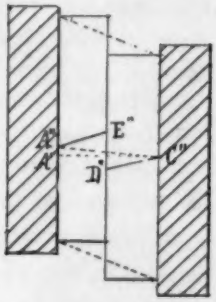


Fig. 3.

two parts, $A''E''$ and $D''C''$; the slip, $D''E''$, being equal to the total displacement of the left side. Two characteristics of the movement are to be noted; the total slip on the ruptured surface equalled the total relative displacement of the blocks of wood; and, at the time of the slip the blocks remain stationary, and the whole movement at that time was an elastic rebound of the jelly to a condition of no strain.

These experiments illustrate as well as simple experiments could what occurred at the time of the California earthquake of April 18, 1906. Fortunately, early surveys had been made of this region which Dr. Hayford, in the report of the California Earthquake Commission has, for the sake of discussion, divided into two groups; I., the surveys made from 1851–65; II., those from 1874–92. A third survey (III.) was made after the earthquake in 1906–7. These surveys extended from Mt. Diablo, about 33 miles east of the fault, to Farallon Light House, about 22 miles west of it. They showed that between the I. and II. surveys Farallon Light House had shifted relatively to Mt. Diablo, 4.6 feet north-northwest, practically in a direction parallel with the fault-line; and between II. and III. sur-

veys it had shifted 5.8 feet more in nearly the same direction, making a total shift in about 50 years of 10.4 feet.

Observations in the field on the offsets of fences and roads showed that at the time of the earthquake there was a relative movement of the two sides at the fault-surface, amounting to something like 20 feet, and it is only reasonable to suppose that this movement was equally divided between the opposite sides of the fault. The surveys show that the actual displacement which took place between II. and III. diminished as the distance from the fault became greater; on the east side the displacement practically died out at a distance of four or five miles from the fault, and on the west side the displacement became equal to that of Farallon Light House at about the same distance from the fault. All the phenomena were in close accord with the experiments described above. The main difference consists in the fact that a straight line on the earth's surface across the fault and at right angles to it did not break up into two straight lines, as in the experiment, but into two curved lines. We ascribe this curvature to the fact that the forces which produced the displacement of the ground were applied below the crust of the earth, whereas in the experiment they were applied at the outer boundary of the jelly.

The elastic rebound near the fault-surface, of course, took place suddenly at the time of the earthquake; and the surveys show that between I. and II., and between II. and III. there was a relative shift of very extensive regions on opposite sides of the fault, but the surveys do not determine whether these shifts took place suddenly at the times of the great earthquakes of 1868 and 1906, or whether they were the effect of a slow, gradual movement continuing through the years. We must turn to other considerations to decide this point. In the experiments we have described the elastic rebound was greatest at the ruptured surface, became progressively less at greater distances from this surface, and the jelly in contact with the wooden blocks did not partake of the movement at all. The experiments might have been varied and instead of a slow shift of the block gradually setting up an elastic shear, we might have set up the shear suddenly; but this was not

PROC. AMER. PHIL. SOC. XLVIII. 192 U, PRINTED SEPTEMBER 7, 1909.

necessary to produce the phenomena which we know took place at the time of the earthquake. It seems impossible to think that the general shift was sudden; for we cannot imagine what forces could have produced a sudden displacement, amounting to four or five feet, of a portion of the earth's surface covering thousands of square miles. But we have indubitable evidence, in the foldings of the rock common to all mountain chains, of the slow displacement of large regions to considerable distances; and unless such a displacement were slow enough to allow the rock everywhere to flow viscously and thus adjust itself to its new position, there would be places where the elastic stresses would from time to time be greater than the strength of the rock and ruptures would occur causing earthquakes.

This view of the case is so entirely in accord with the elastic properties of rock, and with the slow movements of large regions, familiar to geologists, that it commends itself strongly without further argument; but there is a consideration which seems almost decisive in its favor. In the experiments described we saw that the relative slip at the ruptured surface was exactly equal to the total relative shift of the wooden blocks; this, of course, was independent of the slow or sudden nature of the shift. The slip on the fault-surface at the time of the California earthquake was about 20 feet; therefore the shift of the more distant regions which brought about the break must have been as great; but the surveys show that between II. and III., the shift was only 5.8 feet, and between I. and II., 4.6 feet; that is, in all, only about 10.4 feet since the earliest surveys, some 50 years before the shock. We can therefore say, definitely, that the shift which set up the elastic strains which finally resulted in the earthquake, not only did not wholly take place at the time of the rupture but that even fifty years earlier it had already accumulated to about one half its final amount; that between the I. and II. surveys it increased to about three-quarters of this amount, and that the last quarter was added between the II. and III. surveys. It is hardly possible, in view of this history not to be convinced that the shift accumulated gradually.

Since the general order of events, that is, the setting up of elastic strains resulting in the rupture of the rocks which preceded

and caused the California earthquake, were the consequences not of special conditions but of the general properties of rock, we may make the general statement that *tectonic earthquakes are caused by the gradual relative displacement of neighboring regions, which sets up elastic strains so great that the rock is ruptured; and that at the time of the rupture no displacements of large areas take place, but there occurs merely an elastic rebound, to an unstrained position, of the lips of the fault extending but a few miles on each side of it.*

It is not necessary of course that the slow displacement should set up a simple horizontal shear, as in the case of the California earthquake, but simply that an elastic strain of some kind should be produced by the relative displacement of adjoining regions. This may be due, for instance, to the slow sinking of a large region with the production of vertical elastic shears around its boundary, and when these shears become sufficiently strong a break will occur and the movement of the two lips will be vertical and in opposite directions, thus producing a fault-scarp. The main, sinking region, however, would not suddenly drop at the time of the break; there would only be an elastic rebound around its boundaries; its own displacement having taken place slowly over a long period of time. The elastic strains might also be set up by a horizontal compression, in which case the rock would be folded upward, and when the curvature became too great it would break like a bent stick, both sides of the broken surface flying upwards under the elastic forces and leaving an open fissure between them. Examples of this kind of rupture are only known on a small scale.

It is possible that the rupture may not be confined to a single surface, but may be distributed over a number of neighboring surfaces, and a small block between these surfaces may be displaced as a whole; but this must be looked upon as a minor phenomenon of the fault-zone, and is not an example of the readjustment of large blocks.

(*b*) Some Characteristics of Seismological Instruments.

When efforts began to be made, some thirty or forty years ago, to produce an instrument that would record the actual movement of the ground caused by an earthquake, the object aimed at was to

produce a "steady mass," that is, a heavy mass that would remain at rest in spite of the movement of its support; and by recording, either directly or through magnifying levers, its movement relative to the ground, the hope was entertained that the actual movement of the ground would be obtained. But the hope was futile. Every seismograph consists essentially of two parts: a heavy mass adjusted in a greater or less degree to a condition of neutral equilibrium, and the drum or other surface on which the record is made. If the mass could be adjusted absolutely to neutral equilibrium and could be kept in that condition in spite of the movement of its support, it would remain at rest, and would record the true movement of the earth; but the size of the recording apparatus is limited and in order that the record should be made on it, the heavy mass must remain pretty closely in one position, which is practically incompatible with neutral equilibrium. It was found necessary to keep the mass in stable equilibrium although the force brought into play by a small displacement might be very small. If displaced the mass would, therefore, vibrate about its position of equilibrium with a period of its own; and the record of every earthquake is the combination of the earth's movement with that of the heavy mass; and if the period of the vibrations of the earth happens to approach that of the heavy mass, the amplitude of the latter increases greatly, and indicates a movement of the earth much larger than actually occurs. We cannot deduce the movement of the earth from the record except by a careful analysis based on the mathematical theory of the seismograph. This, fortunately, has been worked out; but, unfortunately, it is rather complicated, and it is only in comparatively simple cases that it can be applied without very great labor.

The earlier investigators also thought that all solid friction or viscous damping reduced the sensitiveness of the instrument, and that a long period of vibration increased it. Solid friction is indeed always harmful and should be reduced as much as possible, but viscous damping is a great advantage and simplifies the interpretation of the record. Remembering that every earthquake consists of vibrations of many periods, a glance at figure 4 will show the great benefit of strong damping. The curves show the magnifying power of the seismograph so far as it depends upon the ratio of the

period of the earth's vibration to that of the seismograph itself, and upon the viscous damping. The damping ratio is the ratio of the amplitude of successive swings of the heavy mass, when it is

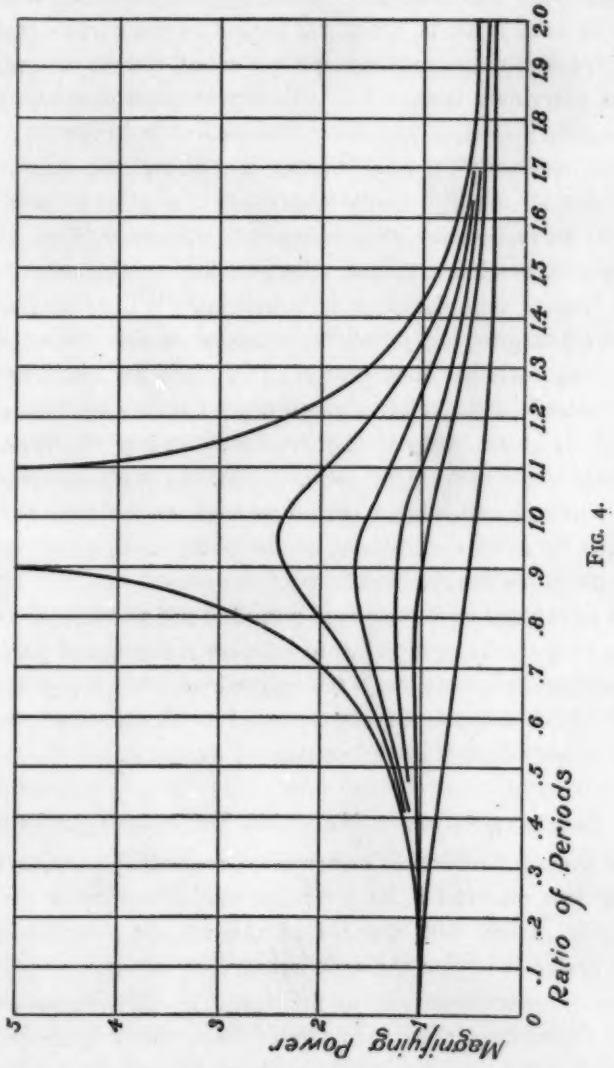


FIG. 4.

allowed to swing freely. If this ratio is nearly 1 : 1, that is, if there is very little damping and the amplitude of the swinging mass dies

down very slowly, the curves show that the magnifying power for vibrations of very short period is unity; that is, the record gives the true amplitude of the earth's motion; for vibrations of longer period the magnifying power rapidly increases, and when the ratio of the periods is unity; that is, when the period of the earth's motion and the free period of the seismograph are equal, the magnifying power becomes extremely large. For still longer periods the magnifying power again decreases and when the period becomes very long, it becomes extremely small. Since, therefore, the vibrations of various periods are differently magnified, it is quite evident that the record of an earthquake would be greatly distorted, some vibrations being unduly emphasized, and others unduly minimized. It is just in this respect that damping is beneficial. Within limits, the inequality of magnifying power for various periods becomes less as the damping ratio becomes greater; and when the damping is great enough to reduce the relative amplitude of successive swings in the ratio of 8:1, the magnifying power is nearly uniform for all periods less than that of the seismograph. A seismograph, damped to this amount, and with a period as long as the longest of those present in the earth's vibrations, would give a much truer representation of the earth's movement.

The advantage of a long free period is not to increase the sensitiveness of the seismograph but to increase the range of periods over which its sensitiveness may be maintained. Contrary to a very general belief, the magnifying power for vibrations of very short periods is not affected by the amount of damping.

(*c*) Suggestions for a National Seismological Bureau.

The work of collecting information regarding earthquakes, and studying this material is so extensive that it cannot be carried out thoroughly except with the aid of the federal government. The United States is almost the only country of importance which does not give governmental aid to the study of earthquakes; and, although, fortunately, the larger part of this country is only subject to occasional slight shocks, extremely destructive shocks have occurred within our boundaries, and certain districts are frequently visited by earthquakes which cause much damage. The study of earth-

quakes is a thoroughly practical subject, and if properly prosecuted, will be of distinct benefit to the country.

Let us glance, for a moment, at the special problems which a national bureau should take up. They may be enumerated as follows:

1. The collection of information regarding earthquakes in the United States and its possessions.

2. The study of the distribution of earthquakes in the United States and the preparation of maps showing this distribution and its relation to the geological structure.

3. The study of special regions, such as the California coast.

4. The prompt examination of a region which has suffered a severe earthquake.

5. The collection of information regarding earthquakes under the sea, and tidal waves.

6. The study of the earthquakes of the Gulf of Mexico and the Caribbean Sea from the records of instruments around these areas.

7. The issue of monthly bulletins, giving the records of felt earthquakes and of seismographs in the United States.

8. The study and dissemination of information regarding the best methods of construction in areas subject to earthquakes.

9. The theoretical study of earthquake instruments.

10. Other theoretical studies.

The variety of these studies requires the sympathetic cooperation of many branches of the government for their successful prosecution. The Weather Bureau and the Post Office Department are especially adapted to collect information regarding felt earthquakes; and the trained observers of the former, distributed as they are all over the country, could readily add a seismograph to the instruments under their charge and obtain important records of distant and near earthquakes. The Navy, through its personnel and through its Hydrographic Office has especial facilities for collecting information regarding earthquakes felt at sea. The Geological Survey alone could study the relation of geological structure to the occurrence of earthquakes; and the Coast and Geodetic Survey has on its staff able mathematicians capable of deducing the characteristics of the interior of the earth from the velocity of earthquake waves through

it, and of finding the answer to the question whether earthquakes produce changes in the earth's magnetism.

In looking over the history of the various scientific bureaus of the government, we see that they were, in general, started by the Smithsonian Institution, and after their work had been thoroughly marked out and justified, they became independent. It seems not only conservative, but most practical, to follow this precedent in the establishment of a seismological bureau; for the Smithsonian is excellently adapted for prosecuting earthquake studies, and it could probably secure the hearty cooperation of all the other departments of the government more easily than could any single one of these departments.

SOME BURIAL CUSTOMS OF THE AUSTRALIAN ABORIGINES.

By R. H. MATHEWS, L.S.

(*Read May 21, 1909.*)

Oval-shaped objects used in connection with native burials in the valley of the Darling River, New South Wales, were manufactured from burnt gypsum,[1] reduced to a powder, and fine sand or ashes, well compounded with water, just as we would mould anything of the kind out of cement or plaster of paris. The necessary shape could be given to the mass while plastic and then allowing it to dry in the sun. These objects are in the shape of a large egg, varying in length from about three to nine inches, by a width of say two and a quarter inches for the smaller ones, up to double that width for the larger. (See Figs. 1, 2, 3 and 4, page 314.)

They are often approximately circular in a section through the middle part, but in other cases such a section would be ovate. Some of them are flattish on one or both sides and are not unlike a cake baked in an elongated form. In a few of the flattened productions, one side is slightly concave, but whether this was intended by the maker it is difficult to say. Probably the wet mass assumed this shape when drying in the sun, because the heat would naturally cause the outer margin, which would dry first, to turn upward, similarly to the way a board warps toward the sun, when exposed in a free state. Nearly all the specimens I have seen were evidently manufactured in the way above described, but an occasional one consists of a piece of sandstone or shale, of a light color, found in the bush, which required but little fashioning to bring it to the required shape.

An old aboriginal, of the Ngunnhalgu tribe, known as Harry Perry by the white people, told me that these *kopai* objects, which he

[1] Called *kopai* by the natives; often erroneously written *copi* and *kopi* by the European residents of that region.

called *mūrndu,* were made out of powdered kopai and a little sand or ashes, much in the way we mix up flour when making dough for baking into bread. He said that when a native of either sex died and was buried, the relatives came to the grave and placed these

1 2 3 4

This picture shows three medium sized cakes and one small one, all of which are made from gypsum (*kopai*), as above described. I shall call them *murndu,* their native name in the Ngunnhalgu tribe, which occupied the country from about Wilcannia up to near Louth, being the tract from various parts of which my specimens were obtained.

FIG. 1. The *murndu* numbered 1 in the picture, is 6¾ inches long, by a maximum width of 4¾ inches. The thickest part, at right angles to the width, is 3⅞ inches. The weight of the article is 2 lbs. 9 oz.

FIG. 2 measures 2¾ inches in length, by a mean thickness of 2⅛ inches. Weight, 4½ oz.

FIG. 3 has a length of a little over 7⅞ inches and its greatest breadth is 4⅛ inches. It is oval in section, with a thickness of 3¼ inches. Weight, 2 lbs. 14 oz.

FIG. 4 is 6⅝ inches in length, with a maximum breadth of 3⅛ inches. It has a practically circular section through the middle. Weight, 2 lbs. 8 oz.

Scattered here and there through the composition of the balls are pieces of gypsum as large as gravel, showing that the mineral was not very well pulverized; a fact which does not surprise us, when we remember that the natives had to burn the gypsum in a camp fire. For the same reason the powder became mixed with small quantities of wood ashes.

kopai balls on top of the mound of earth. For example, if the body were that of an adult man, his widow would place a *mūrndu* on the ground above his head. The deceased's brothers would each place one or more along one side of the grave; his mother and sisters might also lay a *mūrndu* or two on the other side; and so on.

An old man of the Murawarri tribe informed me that in his language the *kopai* ball or tablet is called *yūrda*. When a man, woman, or young person beyond the age of childhood, died, leaves were strewn over the earth covering the grave, and on top of the leaves were laid the *yūrda*. There might be only one or two *yūrda* deposited, or there might be more, depending upon whether the deceased had few or many friends. Mr. E. J. Suttor tells me that he has seen a dozen or more of these kopai balls lying on a native's grave. They were put on as soon as the corpse was buried.

A Ngēumba blackfellow told me that in his tribe the name of the *kopai* balls is *dhaura*. The gypsum was collected, burnt and pounded fine by the women, and the men shaped the *dhaura*.

A resident informs me that gypsum is very plentiful on Yantara Station, near Lake Cobham, about 120 miles northwesterly from the Darling River, where tons of it could easily be obtained. Another correspondent, at Kallara Station on the Darling, states that gypsum is quite plentiful there. In fact, gypsum and pipeclay are both easily obtainable along the valley of the Darling, as well as in the hinterland, all the way from its junction with the Murray River up to Brewarrina. There is also a kind of slacked or rotted gypsum which occurs in patches, resembling slacked lime.

Old Perry and others above quoted said that the object of decorating the grave in the way described was to induce the *bo-ri* or spirit of the dead person, to remain in its place of sepulture and thus prevent its roaming through the camp at night to do injury to anyone with whom the deceased might in his or her lifetime have had a feud. When the spirit saw that its owner's death had been properly mourned for in accordance with the tribal custom, it felt more friendly towards everybody. The spirit comes up during the night and sits on top of the grave and commences licking or sucking one or more of the kopai balls.

Sir Thomas L. Mitchell is the first author to mention these *kopai* balls. He says:

> It was on the summit of a sandhill where I fixed my depot on the Darling [Fort Bourke] that we saw the numerous white balls, and so many graves. The balls are shaped as in the accompanying woodcut, and were made of lime. . . . A native explained one day to Mr. Larmer [a member of Sir

Thomas's Staff] in a very simple manner the meaning of the white balls, by taking a small piece of wood, laying it in the ground and covering it with earth. Then laying his head on one side and closing his eyes, he showed that a dead body was laid in that position in the earth, where these balls were placed above.[3]

In 1901, Mr. G. Officer, of Kallara Station, described some kopai balls or cakes found at a grave on Curronyalpa run on the Darling River, about fifteen miles above Tilpa. There were thirty-nine specimens at the grave, some of which were lying on the surface, others were partially revealed, and the remainder were found by digging a little way into the sandy soil underneath.

FIG. 5 is an exterior view of a *kurno* or widow's cap, *a* being the front, or part fitting over the forehead, whilst *b* represents the back of the head.

Owing to the unusually large number of pieces on this grave, I am inclined to believe that the greater portion of them had been carried from other graves in the neighborhood to this spot and hidden, for the purpose of protecting them from the vandalism of the white men, who were in the habit of carrying them away as curios. Mr. Higgins, a long resident of the Darling region, writes me that two old blackfellows had stated to him that, when the natives observed that the white people desecrated their burying places in this way, they themselves buried the kopai balls in the ground to keep them

[3] "Three Expeditions into Eastern Australia" (London, 1838), Vol. I., pp. 253–4. Seven *kopai* balls are illustrated in the woodcut referred to.

out of sight. Possibly nearly all the specimens recovered by Mr. Officer had originally been concealed with earth, but the violent winds of that district had blown the sandy soil away and left them visible. The grave was on a sandhill about three miles back from the river and was therefore out of the way of the white men, whose principal traffic lay along the course of the stream.

Helmet-shaped objects, called *kurno,* known to have been worn on the heads of widows as a sign of mourning, were made from gypsum, burnt and pounded fine, and mixed with water. A fiber

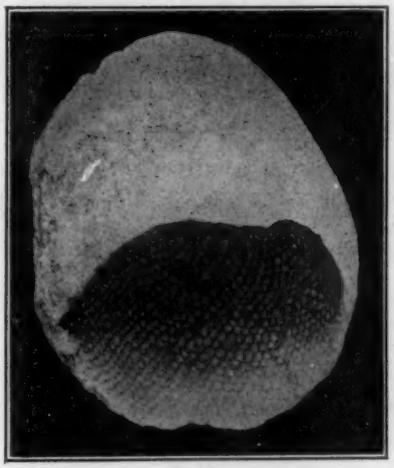


FIG. 6 shows the interior of the cap, with the marks or impression of the net, and the size of its meshes, plainly discernible. This cap weighs 11 lbs. 1 oz., and has been formed of *kopai* or gypsum in the way already described. The specimen was found on a native grave on Lower Budda run, Darling river. I am indebted to Mr. F. W. Beattie for the two photographs, which he took at my request.

or rush net was first placed on the woman's head to protect the hair, and the soft mixture applied outside until it resembled a cap, hence called "widow's caps" by the Europeans. The mixture was not all put on at the same time but by a series of additions extending over a few weeks. The marks of the meshes of the net are distinctly visible in the interior of some of the "caps" of this kind which have been preserved by white men. When the mourning cap had been worn

the customary time, it was taken off and placed by the widow upon the grave of her late husband. When the deceased left a plurality of widows, each wore an emblem of mourning and disposed of it in the same way. If the net was firmly embedded in the dried gypsum, it was left in it, but if the net could be readily detached it was taken out of the cap for future use. In some cases, portions of the woman's hair had to be cut to get the cap off. If the net was left in the cap, it rotted away, but its impression remained. (See Figs. 5 and 6, pages 316 and 317.)

Sir Thomas L. Mitchell reports that on the Darling River he found "Casts in lime or gypsum, which had evidently been taken from a head, the hair of which had been confined by a net, as the impression of it, and some hairs, remained inside." The same author states that, on the Murray, some distance above its confluence with the Darling, he saw some native graves with mounds of earth raised over them, on which were laid the "singular casts of the head in white plaster" which he had before seen at Fort Burke. In some cases the casts of the head were found lying beside the gypsum balls. He gives illustrations of these two "casts," showing also the marks of the net inside.[3]

In 1838, Mr. Joseph Hawdon observed some skull-shaped caps, made of white plaster, which he thought was obtained by burning shells and grinding them into powder. They were laid on the grave of a native near Lake Bonnie on the Murray River. He says that inside the cap was a network of twine. Mr. Hawdon states that he also noticed a great quantity of crystallized lime or gypsum in the locality; it was in masses some tons weight.[4]

Mr. E. J. Eyre gives an example of the "Korno, or widow's mourning cap, made of carbonate of lime, moulded to the head.' The specimen illustrated by him weighed 8½ lbs.[5]

[3] *Op. cit.*, Vol. I., pp. 253–254, and Vol. II., p. 113.

[4] "Diary of an Overland Journey from Port Phillip to Adelaide in 1838" (MSS).

[5] "Journs. Expeds. Discov. Cent. Australia" (London, 1845), Vol. II., p. 509, Plate I., Fig. 17.